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MATHÉMATIQUES

SUR LA REPRÉSENTATION DES FONCTIONS PAR DES FORMULES ANALOGUES À LA FORMULE DE FOURIER

Par J. NATANSON

(Présenté par S. N. Bernstein, de l'Académie, le 3 XI 1935)

1°. Dans cette note je vais indiquer une méthode générale pour construire des formules analogues à celle de Fourier.

Soit $f(t)$ une fonction représentable sous la forme

$$f(x) = \lim_{n\to\infty} \int_a^b f(t)\,\Phi(t, x;\ n)\,dt, \tag{1}$$

où

$$\Phi(t, x;\ n) = \int_\gamma^n K(t, x, z)\,dz. \tag{2}$$

En substituant cette expression dans l'égalité (1), et en intervertissant l'ordre des intégrations on obtient la formule:

$$f(x) = \int_\gamma^\infty dz \int_a^b f(t)\,K(t, x, z)\,dt. \tag{3}$$

C'est la formule que nous cherchions.

2°. Soit, par exemple, $f(t)$ une fonction représentable par l'intégrale de Dirichlet:

$$f(x) = \lim_{n\to\infty} \frac{1}{2\pi} \int_0^{2\pi} f(t)\,\frac{\sin\dfrac{2n+1}{2}(t-x)}{\sin\dfrac{t-x}{2}}\,dt, \tag{4}$$

or

$$\sin\left(n+\frac{1}{2}\right)(t-x) = \int_0^{n+1/2} (t-x)\cos z(t-x)\,dz,$$

donc

$$f(x) = \frac{1}{2\pi} \int_0^\infty dz \int_0^{2\pi} f(t)\cdot\frac{t-x}{\sin\dfrac{t-x}{2}}\cdot\cos z(t-x)\,dt. \tag{5}$$

3°. Comme un autre exemple considérons l'intégrale de M. Fejér:

$$f(x) = \lim_{n\to\infty} \frac{1}{2n\pi} \int_0^{2\pi} f(t) \left(\frac{\sin n \frac{t-x}{2}}{\sin \frac{t-x}{2}} \right)^2 dt. \tag{6}$$

En transformant cette relation, on obtient:

$$f(x) = \frac{1}{4\pi} \int_0^{\infty} \frac{dz}{z^2} \int_0^{2\pi} f(t) \frac{\sin z(t-x)}{\sin^2 \frac{t-x}{2}} \left[z(t-x) - \operatorname{tg} \frac{z}{2} (t-x) \right] dt. \tag{7}$$

Cette relation subsitse pour toutes les fonctions sommables $f(t)$ telles que l'on a:

$$\lim_{n\to 0} \frac{1}{n} \int_0^n |f(t+x) - f(x)|\, dt = 0. \tag{8}$$

4°. En appliquant cette transformation à la formule de Poisson

$$f(x) = \lim_{r\to 1} \frac{1}{2\pi} \int_0^{2\pi} f(t) \frac{1-r^2}{1-2r\cos(t-x)+r^2}\, dt, \tag{9}$$

on en tire immédiatement

$$f(x) = \frac{1}{\pi} \int_0^1 dz \int_0^{2\pi} f(t) \left\{ \varphi(z) + \frac{\cos(t-x) + z^2 \cos(t-x) - 2z}{[1-2z\cos(t-x)+z^2]^2} \right\} dt, \tag{10}$$

où $\varphi(z)$ est une fonction arbitraire, telle que l'on a:

$$\int_0^1 \varphi(z)\, dz = \frac{1}{2}.$$

Cette formule est valable pour n'importe quelle fonction $f(t)$ sommable et telle que l'on ait:

$$\lim_{n\to 0} \frac{1}{n} \int_0^n \left\{ f(t+x) - f(x) \right\} dt = 0.$$

5°. Les formules bien connues de Riemman-Lebesgue:

$$\lim_{n\to 0} \int_a^b f(t) \sin nt\, dt = \lim_{n\to\infty} \int_a^b f(t) \cos n\, t\, dt = 0$$

peuvent être représentées sous la forme suivante:

$$\int_0^{\infty} dz \int_a^b f(t) \cos z\, t\, dt = 0; \int_0^{\infty} dz \int_a^b f(t)\, t \sin z\, t\, dt = \int_a^b f(t)\, dt. \tag{11}$$

6°. Tous les théorèmes de la théorie des intégrales singulières se transforment ainsi en des théorèmes portant sur les formules du type (3). Considérons, par exemple, les trois théorèmes suivants de M. Lebesgue:

Théorème A. Pour que la relation (1) ait lieu pour n'importe quelle fonction $f(t)$ sommable et continue en le point x, il faut et il suffit que le noyau $\Phi(t, x; n)$ satisfasse aux conditions suivantes:

1) Si $\alpha < x < \beta$, on a $\lim_{n\to\infty} \int_{\alpha}^{\beta} \Phi(t, x; n)\, dt = 1$.

2) $\int_{\alpha}^{b} |\Phi(t, x; n)|\, dt \leqslant K(x)$.

3) Si $x < \alpha$ ou $x > \beta$, on a (pour $\alpha \leqslant t \leqslant \beta$) $|\Phi(t, x; n)| \leqslant L(x; \alpha, \beta)$.

Théorème B. Pour que la relation (1) ait lieu pour n'importe quelle fonction $f(t)$ à variation bornée et continue en le point x, il faut et il suffit que le noyau $\Phi(t, x, n)$ satisfasse aux conditions suivantes:

1) Si $\alpha < x < \beta$, on a $\lim_{n\to\infty} \int_{\alpha}^{\beta} \Phi(t, x; n)\, dt = 1$.

2) Pour toutes les valeurs de n, α, β on a;

$$\left| \int_{\alpha}^{\beta} \Phi(t, x; n)\, dt \right| \leqslant K(x).$$

Théorème C. Pour que la relation (1) ait lieu pour n'importe quelle fonction $f(t)$ (mesurable et bornée) approximativement continue en le point x, il faut et il suffit que le noyau $\Phi(t, x; n)$ satisfasse à la condition

$$\lim_{n\to\infty} \int_{E} \Phi(t, x; n)\, dt = 1,$$

quel que soit l'ensemble E de densité 1 au point x.

On peut énoncer ces théorèmes dans les formes suivantes*:

Théorème A*. Pour que la relation (3) ait lieu pour n'importe quelle fonction $f(t)$ sommable et continue en le point x, il faut et il suffit que le noyau $K(t, x, z)$ satisfasse aux conditions suivantes:

1) Si $\alpha < x < \beta$, on a $\int_{l}^{\infty} dz \int_{\alpha}^{\beta} K(t, x, z)\, dt = 1$.

2) $\int_{a}^{b} dt \left| \int_{l}^{n} K(t, x, z) \right| dz \leqslant K(x)$.

3) Si $x < \alpha$ ou $x > \beta$, on a

$$\left| \int_{l}^{n} K(t, x, z)\, dz \right| \leqslant L(x; \alpha, \beta). \quad (\alpha \leqslant t \leqslant \beta)$$

Théorème B*. Pour que la relation (3) ait lieu pour n'importe quelle fonction $f(t)$ à variation bornée et continue en le point x, il faut et il suffit que le noyau $K(t, x, z)$ satisfasse aux conditions suivantes:

1) Si $\alpha < x < \beta$, on a $\int_{l}^{\infty} dz \int_{\alpha}^{\beta} K(t, x, z)\, dt = 1$.

2) $\left| \int_{l}^{n} dz \int_{\alpha}^{\beta} K(t, x, z)\, dt \right| \leqslant K(x)$, pour toutes les valeurs de n, α, β.

---

* En admettant que l'interversion de l'ordre des intégrations soit légitime.

Théorème C*. Pour que la relation (3) ait lieu pour n'importe quelle fonction $f(t)$ (mesurable et bornée) approximativement continue en le point x, il faut et il suffit que le noyau $K(t, x, z)$ satisfasse à la condition:

$$\int_1^\infty dz \int_E K(t, x, z)\, dt = 1,$$

quel que soit l'ensemble E de densité 1 au point x.

7°. Exemple. Posons $K(t, x, z) = \frac{1}{\pi} \cos z(t - x)$. Alors ce noyau satisfait aux conditions du théorème B*. En effet, si $\alpha < x < \beta$, nous avons:

$$\int_z^\beta K(t, x, z)\, dt = \frac{1}{\pi}\left[\frac{\sin z(\beta - x)}{z} - \frac{\sin z(\alpha - x)}{z}\right],$$

d'où

$$\int_0^\infty dz \int_\alpha^\beta K(t, x, z)\, dt = \frac{1}{\pi}\int_0^\infty \frac{\sin z(\beta - x)}{z}\, dz - \frac{1}{\pi}\int_0^\infty \frac{\sin z(\alpha - x)}{z}\, dz = 1.$$

Il est aisé de vérifier la condition 2) du théorème B*. Alors, quelle que soit la fonction $f(t)$ à variation bornée et continue en le point x, nous avons:

$$f(x) = \frac{1}{\pi}\int_0^\infty dz \int_a^b f(t) \cos z(t - x)\, dt.$$

C'est la formule de Fourier.

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MATHÉMATIQUES

SUR LES ESPACES MÉTRIQUES D'UN CERTAIN TYPE

Par B. VULICH

(Présenté par S. N. Bernstein, de l'Académie, le 23 XI 1935)

Dans cette note, développant les idées du prof. L. Kantorovitch, dont il m'a obligeammant donné communication, je construis un nouveau type de métrique d'un espace linéaire, permettant d'obtenir comme convergence d' une suite d'éléments de cet espace un mode de convergence tel, qu'il ne peut s'obtenir à l'aide des espaces normés au sens de M. S. Banach (1). Les détails de mes recherches et la démonstration des résultats présentés ci-dessous seront publiés ailleurs.

1°. Soit X un espace linéaire d'éléments x et tel qu'à chaque complexe K composé d'un nombre fini d'éléments $x_1, x_2, \ldots, x_\nu$, que nous désignons par $K = (x_1, \ldots, x_\nu)$, corresponde un nombre non négatif

$$\| K \| \text{ ou } \| (x_1, \ldots, x_\nu) \|,$$

appelé — n o r m e de ce complexe. Soit donné que ce nombre ne dépende pas de l'ordre de numérotage de éléments du complexe envisagé, et satisfasse aux axiomes suivants:

I. Si un ou plusieurs éléments du complexe K y entrent plusieurs fois (le nombre des répétitions pouvant différer d'un élément à l'autre), le complexe K_1 composé de tous les divers éléments du complexe K, mais n'y entrant chacun qu'une seule fois, a la même norme.

II. On a $\| x \| = 0 \supset x = 0$, où $\| x \|$ signifie la norme du complexe composé d'un seul élément x.

III. Si K est un complexe composé de tous les éléments entrant dans les complexes K_1 et K_2, alors on a

$$\| K \| \leqslant \| K_1 \| + \| K_2 \|.$$

IV. $\| (x_1, \ldots, x_\nu) \| \leqslant \| (x_1, \ldots, x_\nu, x_{\nu+1}) \|$.

V. $\| (x_1^{(1)} + x_1^{(2)}, \ldots, x_\nu^{(1)} + x_\nu^{(2)}) \| \leqslant$

$$\leqslant \| (x_1^{(1)}, \ldots, x_\nu^{(1)}) \| + \| (x_1^{(2)}, \ldots, x_\nu^{(2)}) \|.$$

VI. $\alpha_1, \ldots, \alpha_\nu$ étant des nombres réels, on a

$$\| (\alpha_1 x_1, \ldots, \alpha_\nu x_\nu) \| \leqslant \max_{1 \leqslant i \leqslant \nu} | \alpha_i | \cdot \| (x_1, \ldots, x_\nu) \|.$$

Nous dirons n o r m é K l'espace linéaire X, dans lequel la notion de norme du complexe satisfaisant aux axiomes énoncés est définie.

2°. Nous dirons convergente K vers x_0 une suite $\{x_n\}$ si quel que soit $\varepsilon > 0$, il existe un $N > 0$ tel, que pour chaque $n \geqslant N$ et chaque $p \geqslant 0$ on ait

$$\|(x_n - x_0, \ldots, x_{n+p} - x_0)\| < \varepsilon,$$

et nous désignerons cette propriété de la suite par l'expression

$$x_n \underset{n\to\infty}{\overset{K}{\longrightarrow}} x_0.$$

On peut considérer un espace X normé K comme un espace normé au sens de M. S. Banach, lorsqu'on prend pour norme de chaque élément de cet espace la norme du complexe composée de cet élément. Si $\{x_n\}$ converge vers x_0 suivant la norme, c.-à-d. que

$$\lim_{n\to\infty} \|x_n - x_0\| = 0,$$

nous écrirons

$$x_n \underset{n\to\infty}{\overset{B}{\longrightarrow}} x_0.$$

Théorème 1. *Si* $x_n \underset{n\to\infty}{\overset{K}{\longrightarrow}} x_0$, *alors on a* $x_n \underset{n\to\infty}{\overset{B}{\longrightarrow}} x_0$.

Théorème 2. *Si* $x_n \underset{n\to\infty}{\overset{B}{\longrightarrow}} x_0$, *il existe une suite partielle* $\{x_{n_i}\}$ *telle, que l'on a*

$$x_{n_i} \underset{i\to\infty}{\overset{K}{\longrightarrow}} x_0.$$

3°. Nous dirons que l'opération additive $y = U(x)$, où $x \in X$, $y \in Y$ et X et Y sont des espaces normés K, est linéaire K, si l'on a toujours

$$x_n \underset{n\to\infty}{\overset{K}{\longrightarrow}} x \quad \supset \quad U(x_n) \underset{n\to\infty}{\overset{K}{\longrightarrow}} U(x);$$

De façon analogue nous dirons linéaire K une fonctionnelle additive $F(x)$, si l'on a toujours

$$x_n \underset{n\to\infty}{\overset{K}{\longrightarrow}} x \quad \supset \quad F(x_n) \underset{n\to\infty}{\longrightarrow} F(x);$$

Nous entendrons par opération (resp. fonctionnelle) linéaire B l'opérations (resp. fonctionnelle) linéaire au sens de S. Banach.

Théorème 3. *Les notions de fonctionnelles linéaires K et linéaires B sont équivalentes, tandis que l'opération linéaire K est un cas particulier de l'opération linéaire B.*

Il existe donc une opération linéaire B, qui n'est pas en même temps une opération linéaire K.

4°. $y = U(x)$ étant une opération définie dans l'espace X, avec son contredomaine situé dans l'espace Y, si $\mathrm{K} = (x_1, \ldots, x_\nu)$, nous désignerons par $U(\mathrm{K})$ le complexe des éléments de l'espace Y

$$U(K) = \Big(U(x_1), \ldots, U(x_\nu)\Big);$$

Théorème 4. *Pour qu'une opération additive $U(x)$ soit linéaire K, il faut et il suffit qu'il existe un nombre $C \geqslant 0$ tel que pour tout complexe* K *l'on ait*

$$\|U(\mathrm{K})\| \leqslant C\|\mathrm{K}\|.$$

5°. Considérons des espaces $\mathfrak{R}_X$ et $\mathfrak{R}_Y$ de suites d'éléments appartenant respectivement à X et Y convergentes K vers 0. On peut définir une norme pour chaque élément de $\mathfrak{R}_X$ en posant

$$\| \{x_n\} \| = \lim_{\nu \to \infty} \| (x_1, \ldots, x_\nu) \|$$

et d'une façon analogue pour l'espace $\mathfrak{R}_Y$.

Faisons correspondre à l'aide d'une opération linéaire K $y = U(x)$ à chaque élément de $\mathfrak{R}_X$ un élément de $\mathfrak{R}_Y$. On voit facilement qu'en vertu du théorème 4 cette correspondance est une opération linéaire B dans l'espace $\mathfrak{R}_X$.

Ainsi donc chaque opération lineaire K peut être considerée comme une opération linéaire B dans un espace plut compliqué et normé seulement au sens de S. Banach.

6°. Exemples d'espaces normés K.

A. L'espace $(L^{(r)})$, où $r \geqslant 1$, des fonctions $x(t)$ avec la $r$ème puissance sommable dans l'intervalle $0 \leqslant t \leqslant 1$.

Posons

$$\| (x_1, \ldots, x_\nu) \| = \left\{ \int_0^1 \left[\max \left(\left| x_1(t) \right|, \ldots, \left| x_\nu(t) \right| \right) \right]^r dt \right\}^{\frac{1}{r}}.$$

Théorème 5. *Pour qu'une suite* $\{x_n\}$ *converge* K *vers* x_0, *où* $x_n \in (L^{(r)})$, $x_0 \in (L^{(r)})$, *il faut est il suffit que l'on ait: a)* $x_n(t) \underset{n \to \infty}{\longrightarrow} x_0(t)$ *presque partout, b) pour tout* $n = 1, 2, \ldots$

$$|x_n(t)| \leqslant \xi(t), \text{ où } \xi(t) \in (L^{(r)}).$$

B. L'espace $(l^{(r)})$, où $r \geqslant 1$, des suites de nombres $x = \{\xi_i\}$ telles que l'on a $\sum_{i=1}^{\infty} |\xi_i|^r < \infty$.

Posons, pour $x_k = \{\xi_i^{(k)}\}$, $(k = 1, 2, \ldots, \nu)$,

$$\| (x_1, \ldots, x_\nu) \| = \left\{ \sum_{i=1}^{\infty} [\max (|\xi_i^{(1)}|, \ldots, |\xi_i^{(\nu)}|)]^r \right\}^{\frac{1}{r}}.$$

7°. L'examen des suites d'opérations définies dans un espace normé K, appliqué à l'espace $(L^{(r)})$ nous amène au résultat suivant:

Soient $\{x_{x_i}^{(k)}(t)\}$ $(i, k = 1, 2, \ldots)$ des fonctions entrant dans $(L^{(r)})$, et soit

(a) pour chaque $i = 1,2, \ldots$ $x_i^{(k)}(t) \underset{k \to \infty}{\longrightarrow} x_i(t)$ presque partout et $|x_i^{(k)}(t)| \leqslant \xi_i(t)$ $(k = 1, 2, \ldots)$, où $\xi_i(t) \in (L^{(r)})$,

(b) pour chaque $k = 1, 2, \ldots$ $\{x^{(k)}(t)\}_i$ converge en moyenne avec la $r$ème puissance vers $x^{(k)}(t)$,

(c) $\{x_i(t)\}$ converge en moyenne avec la $r$ème puissance vers $x(t)$.

Introduisons les notations

$$x_i^{(\infty)}(t) = x_i(t); \quad x^{(\infty)}(t) = x(t).$$

Théorème 6. *Pour que les conditions* (a), (b) *et* (c) *entraînent* (d) $x^{(k)}(t) \underset{k \to \infty}{\longrightarrow} x(t)$ *presque partout et*

(e) $|x^{(k)}(t)| \leqslant \xi(t)$ $(k = 1, 2, \ldots)$, *où* $\xi(t) \in (L^{(r)})$,
il faut et il suffit que, quelque soit $\varepsilon > 0$, *il existe un* $k_0 > 0$ *et un nombre naturel* i *tels que pour chaque ensemble fini de nombres* $k_1, k_2, \ldots, k_\nu$ *satisfaisant à la condition* $k_l \geqslant k_0$ $(l = 1, 2, \ldots \nu)$ *(nous admettons que l'un des nombres* $k_1, \ldots, k_\nu$ *peut prendre la valeur* ∞*) l'on ait*

$$\int_0^1 \max\left(|x_i^{(k_1)}(t) - x^{(k_1)}(t)|^r, \ldots, |x_i^{(k_\nu)}(t) - x^{(k_\nu)}(t)|^r\right) dt < \varepsilon.$$

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LITTÉRATURE CITÉE

[1] S. Banach, Théorie des opérations linéaires, Warszawa, 1932, p. 53.

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MATHÉMATIQUES

SUR LA FONCTION SPECTRALE DE CERTAINES ÉQUATIONS INTÉGRALES HERMITIENNES

Par N. GUNTHER, membre correspondant de l'Académie

1. Convenons de dire 1) que la fonction $u(\tau, x)$ de domaines (τ) formant un corps (C) et de points (x), éventuellement indépendante de (τ) ou de (x), appartient à la classe (U), si a) elle est continue pour les points (x) qui n'appartiennent pas à un ensemble intégrable (E) et ne sont pas situés sur la frontière de (τ); si b) pour chaque (x) n'appartenant pas à (E) elle a pour corps de continuité le corps (C_x), qu'on obtient de (C) en extrayant les domaines (τ), qui ont le point (x) sur leur frontière et si sa variation totale $U(\Omega_y, x)\Omega_y$ est bornée par un nombre B indépendamment de la position de (x); si c) (δ) étant des domaines contenant (E) et contenus dans un domaine (δ_E), on a $U(\delta, x)\delta < \varepsilon$ pour chaque (x) situé en dehors de (δ). La fonction $\delta(\tau, x)$, égale à zéro ou à $\frac{1}{\tau}$, suivant que le point (x) est en dehors de (τ) ou dans son intérieur, appartient à la classe (U). Si $h(x)$ est une fonction arbitraire de classe (U), convenons de désigner, $u(\omega)$ étant une fonction de classe (U) à valeurs positives, par $h(\omega)$ la valeur moyenne mesurée

$$\frac{1}{\omega}\int_{(\omega)} u(\omega)\, h(x)\, d\omega \text{ de } h(x).$$

2. En m'occupant des équations

$$\varphi(x) = \lambda \int_{(\Omega_y)} k(\tau, x)\, \varphi(y)\, d\tau + f(x) \qquad (1)$$

dans lesquelles $k(\tau, x)$ est de classe (U) et $k(\tau, \omega) = \overline{k(\omega, \tau)}$, j'ai donné [1, 2] la valeur moyenne mesurée $r(\tau, \omega, \lambda)$ de la résolvante sous la forme d'une série qui converge pour toutes les valeurs imaginaires de λ et pour les valeurs réelles ne dépassant pas un certain nombre λ en valeur absolue. Le noyau $k(\tau, x)$ était assujetti encore à une condition qui peut être remplacée par l'inégalité

$$\left| \int_{(\Omega_x)} U(\omega)\, \overline{h(x)} \left(\int_{(\Omega_y)} h(y)\, k(\tau, x)\, d\tau \right) d\omega \right| < M \int_{(\Omega_x)} U(\omega) \left| h(x) \right|^2 d\omega, \qquad (2)$$

pour chaque $h(x)$ de classe (U), M étant un certain nombre convenablement choisi; je démontre, toutefois, que cette condition est superflue,

l'inégalité (2) ayant lieu toujours si $M \geqq B$. La construction de la résolvante me permet d'étudier la fonction spectrale de l'équation (1) en me guidant par les idées, exposées dans ma communication au congrès mathématique de Léningrad, 1934.

3. Soit m la mesure de l'intervalle $(\mu^{(1)}, \mu^{(2)})$.

Formons la variable

$$\frac{1}{2\pi i m}\int_{\mu^{(1)}}^{\mu^{(2)}}\left\{\int_{0}^{\mu}\left[r(\tau,\ \omega,\ \mu'+\nu i)-r(\tau,\ \omega,\ \mu'-\nu i)\right]d\mu'\right\}d\mu,\quad 0\leqslant\mu^{(1)},\ \nu>0 \tag{3}$$

en lui faisant changer de signe, lorsque $\mu^{(2)}\leqslant 0$. La variable (3) possède pour chaque valeur choisie de (m) une limite $\vartheta(\tau, \omega, m)$ lorsque $\nu \to 0$. La fonction $\vartheta(\tau, \omega, m)$ est une fonction à variation bornée de (τ) pour (ω) et (m) fixes; on a $\vartheta(\tau, \omega, m) = \overline{\vartheta(\omega, \tau, m)}$; comme fonction de (m) elle est à variation bornée pour (ω) et (τ) fixes absolument continue. Cette dernière propriété montre qu'elle est la moyenne d'une fonction sommable $\vartheta(\tau, \omega, \mu)$:

$$\vartheta(\tau,\ \omega,\ m)=\frac{1}{m}\int_{\mu^{(1)}}^{\mu^{(2)}}\vartheta(\tau,\ \omega,\ \mu)\,d\mu. \tag{4}$$

La fonction $\vartheta(\tau, \omega, \mu)$ comme fonction de μ est bornée; ses points de discontinuité forment un ensemble dénombrable E_0 et ne dépendent pas du choix des domaines (τ) et (ω). On parvient à la démonstration de ces énoncés en partant de l'identité:

$$\frac{r(\tau,\ \omega,\ \lambda_1)-r(\tau,\ \omega,\ \lambda_2)}{\lambda_1-\lambda_2}=\int_{(\Omega_z)}\frac{r(\tau,\ \xi,\ \lambda_1)\,r(\xi,\ \omega,\ \lambda_2)}{u(\xi)}\,d\xi, \tag{5}$$

qui après l'introduction de la fonction $\vartheta(\tau, \omega, \mu)$, si μ et μ_1 n'appartiennent pas à (E_0), prend la forme

$$\vartheta(\tau,\ \omega,\ \mu)=\int_{(\Omega_z)}\frac{\vartheta(\tau,\ \xi,\ \mu)\,\vartheta(\xi,\ \omega,\ \mu_1)}{u(\xi)}\,d\xi,\quad \text{si}\quad 0<\mu<\mu_1; \tag{6}$$

si l'on a $\mu_1<\mu<0$, il faut faire changer le signe de la partie droite de (6).

4. Pour définir la fonction spectrale $\theta(\tau, \omega, m)$, en supposant que $\mu^{(1)}$ et $\mu^{(2)}$ n'appartiennent pas à (E_0), posons

$$\theta(\tau,\ \omega,\ m)\,m=\vartheta(\tau,\ \omega,\ \mu^{(2)})-\vartheta(\tau,\ \omega,\ \mu^{(1)}); \tag{7}$$

la fonction $\theta(\tau, \omega, m)$ est définie dans le corps des intervalles, dont les extrémités n'appartiennent pas à (E_0). L'identité (6) donne immédiatement, si (m) et (m_1) n'ont pas de points intérieurs communs:

$$\int_{(\Omega_z)}\frac{\theta(\tau,\ \xi,\ m)\,\theta(\xi,\ \omega,\ m_1)}{u(\xi)}\,d\xi=0,\qquad \int_{(\Omega_z)}\frac{\theta(\tau,\ \xi,\ m)\,m\cdot\theta(\xi,\ \omega,\ m)\,m}{u(\xi)}\,d\xi=\theta(\tau,\ \omega,\ m)\,m. \tag{8}$$

Eu égard à la première égalité (8), nous dirons que $\theta(\tau, \omega, m)$ satisfait à la condition d'orthogonalité; eu égard à la seconde, nous dirons que $\theta(\tau, \omega, m)$ est normée. Comme fonction de (τ) pour (ω) et (m) fixes, $\theta(\tau, \omega, m)$ est à variation bornée; on a $\theta(\tau, \omega, m) = \overline{\theta(\omega, \tau, m)}$; comme fonction de (m) pour (τ) et (ω) fixes, $\theta(\tau, \omega, m)$ est à variation bornée, la borne supérieure de la variation totale ne dépendant pas de l'étendue de (m). Si $h(x)$ est une fonction arbitraire de classe (U), on a

$$0 < \int_{(\Omega_y)} h(y) \Big(\int_{(\Omega_x)} \overline{h(x)}\, \theta(\tau, \omega, m)\, m\, d\omega \Big)\, d\tau \leqslant \int_{(\Omega_x)} u(\omega)\, |h(x)|^2\, d\omega. \qquad (9)$$

Si l'on pose $\theta(\tau, \omega, m) = \theta_0(\tau, \omega, m) + \theta_1(\tau, \omega, m)$, où $\theta_0(\tau, \omega, m)$ est absolument continue et où les valeurs de $\theta_1(\tau, \omega, m)$ sont presque partout égales à zéro, l'une ou l'autre des fonctions θ_0, θ_1 peut faire défaut; par exemple, si $k(\tau, x) = \delta(\tau, x)$, θ_0 fait défaut, si $k(\tau, x) = \delta(\tau, x) : \omega(x)$, où $\omega(x)$ est constamment croissante, c'est θ_1 qui fait défaut. Si $k_n(\tau, x)$ est un noyau itéré de $k(\tau, x)$, on a

$$\frac{1}{m} \int_{(m)} \frac{\theta(\tau, \omega, m)}{\mu^n}\, dm = \int_{(\Omega_x)} k_n(\tau, z)\, \theta(\xi, \omega, m)\, d\xi. \qquad (10)$$

5. Nous dirons que $\theta(\tau, \omega, m)$ est fermée, si pour chaque fonction $h(x)$ de classe (U) il faut mettre dans (9) le signe d'égalité; si $\theta(\tau, \omega, m)$ n'est pas fermée, nous dirons que $h(x)$ est d'espèce (F), si dans (9) subsiste le signe de l'égalité. Si $h(x)$ est d'espèce (F), nous avons l'égalité

$$h(\omega) = \lim \int_{\mu^{(1)}}^{\mu^{(2)}} \Big(\int_{(\Omega_y)} h(y)\, \theta(\tau, \omega, m)\, d\tau \Big)\, dm, \quad \mu^{(1)} \to -\infty, \quad \mu^{(2)} \to +\infty, \qquad (11)$$

á laquelle donnons la forme de

$$h(\omega) = \int_{(-\infty)}^{(+\infty)} \Big(\int_{(\Omega_y)} h(y)\, \theta(\tau, \omega, m)\, d\tau \Big)\, dm, \qquad (12)$$

signalant par les crochets, que l'intégrale de Stieltjes est obtenue par un passage à la limite, cette limite étant distincte, en général, de l'intégrale prise entre limites infinies (3). On a de même, si $h(x)$ est d'espèce (F) et $h_1(x)$ une fonction arbitraire de classe (U):

$$\int_{(\Delta)} u(\omega)\, h(x)\, h_1(x)\, d\omega = \int_{(-\infty)}^{(+\infty)} \Big(\int_{(\Delta)} h_1(x) \Big(\int_{(\Omega_y)} h(y)\, \theta(\tau, \omega, m)\, d\tau \Big)\, d\omega \Big)\, dm. \qquad (13)$$

Convenant de dire que la fonction moyenne $h(\omega)$ est de classe (H), si l'intégrale

$$\int_{(\Omega_x)} \frac{|h(\omega)|^2}{u(\omega)}\, d\omega \qquad (14)$$

a un sens, on a, si $\theta(\tau, \omega, m)$ est fermée et $h(\omega)$ de classe (H),

$$h(\omega) = \int_{(-\infty)}^{(+\infty)} \Big(\int_{(\Omega_y)} \frac{h(\tau)\, \theta(\tau, \omega, m)}{u(\tau)}\, d\tau \Big)\, d\omega. \qquad (15)$$

6. Désignons par $(m^{(e)})$ l'intervalle complémentaire à (m), c'est-a-dire contenant tous les points de l'axe réel, qui ne sont pas des points intérieurs de (m). En transformant l'équation (1) en l'équation

$$\varphi(x) = \lambda_1 \int_{(\Omega_y)} \Big(\frac{a}{2} k(\tau, x) + \delta(\tau, x) \Big) \varphi(y)\, d\tau + \\ + \frac{a}{2\left(\lambda + \frac{a}{2}\right)} f(x), \qquad \lambda_1 = \frac{\lambda}{\lambda + \frac{a}{2}}, \qquad (16)$$

où a est une constante convenablement choisie, on parvient à définir la valeur de $\theta(\tau, \omega, m)$ pour l'intervalle $(m^{(e)})$ et à démontrer que

$$\theta(\tau, \omega, m^{(e)})\, m^{(e)} + \theta(\tau, \omega, m)\, m = \delta(\tau, \omega). \tag{17}$$

Pour chaque fonction $\theta(\tau, \omega, m)$ et pour toutes les fonctions de classe (U on a

$$h(\omega) = \int_{-\infty}^{+\infty} \Big(\int_{(\Omega_y)} h(y)\, \theta(\tau, \omega, m)\, d\tau \Big)\, dm; \tag{18}$$

$$\int_{(\Delta)} u(\omega)\, h(x)\, h_1(x)\, d\omega = \int_{-\infty}^{+\infty} \Big(\int_{(\Omega_x)} h_1(x) \Big(\int_{(\Omega_y)} h(y)\, \theta(\tau, \omega, m)\, d\tau \Big)\, d\omega \Big)\, dm; \tag{19}$$

si $h(\omega)$ est de classe (H), on a pour chaque $\theta(\tau, \omega, m)$:

$$h(\omega) = \int_{-\infty}^{+\infty} \Big(\int_{(\Omega_y)} \frac{h(\tau)\, \theta(\tau, \omega, m)}{u\tau}\, d\tau \Big)\, dm. \tag{20}$$

En se basant sur (18) on démontre que pour chaque $\theta(\tau, \omega, m)$ on a

$$k_n(\tau, \omega) = \int_{(-\infty)}^{(+\infty)} \frac{\theta(\tau, \omega, m)}{\mu^n}\, dm, \quad r(\tau, \omega, \lambda) = \int_{(-\infty)}^{(+\infty)} \frac{\theta(\tau, \omega, m)}{\mu - \lambda}\, dm. \tag{21}$$

On démontre, enfin, que pour chaque $\theta(\tau, \omega, m)$ la fonction

$$h(x) = \int_{(\Omega_y)} k(\tau, x)\, h_1(y)\, d\tau, \tag{22}$$

où $h_1(x)$ est une fonction de classe (U), est d'espèce (F), d'où suit que pour $h(x)$ subsiste l'égalité (12); c'est le théorème de Hilbert-Schmidt.

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PHYSIQUE

DE L'ÉTABLISSEMENT D'ÉQUIVALENCES ÉLECTROMÉCANIQUES

Par G. GAMBURCEV

(Présenté par P. P. Lasareff, de l'Académie, le 1 VIII 1935)

1. Dans la représentation graphique de systèmes mécaniques il est sous-entendu que le plan du tracé présente une résistance mécanique infiniment petite; par contre, lorsqu'il s'agit de représenter des schémas électriques il est reçu d'admettre que le plan du tracé est caractérisé par une résistance électrique infiniment grande. Cette différence entraine de grandes difficultés pour l'établissement d'équivalences électromécaniques. Toutefois, il est possible, en effectuant la représentation de systèmes mécaniques, d'admettre que le plan du dessin a une résistance mécanique infiniment grande. On en sera quitte pour modifier le procédé de représentation graphique du système mécanique. En justaposant le tracé du système mécanique exécuté par le nouveau procédé avec son analogue électrique, nous verrons que graphiquement ils sont identiques. On peut, d'autre part, changer le procédé de représentation du schéma électrique, en supposant le plan du tracé conducteur idéal; cela nous amènera de même à une analogie ou identité graphique entre schémas équivalents mécaniques et électriques.

Ainsi donc, le problème de l'établissement d'équivalences électromécaniques se trouve ramené à remplacer la procédé habituel de représentation graphique de systèmes mécaniques ou électriques par le nouveau procédé qui vient d'être formulé ci-dessus.

2. Afin qu'à tout schéma électrique il soit possible de mettre en parallèle une analogie mécanique, il est nécessaire de pouvoir prêter aux éléments mécaniques deux pôles (bornes) tout comme aux éléments électriques.

A l'opposé du ressort et de l'amortiseur la masse ne satisfait pas à cette condition, ce qui nous impose certaines réserves dans la poursuite de notre tâche. A titre d'exemple signalons qu'avec une masse du type accoutumé on ne saurait construire une analogie mécanique pour le montage en parallèle de self-inductions. Pour réaliser une analogie de la dite opération on a proposé en qualité d'éléments d'inertie des systèmes formés d'une combinaison de masses avec des leviers ou poulies ([1,2]). En pareil cas la masse, tout comme les autres éléments mécaniques, possède deux «bornes». Toutefois, si nous considérons, par exemple, l'action de l'élément proposé par Lehr ([1]) et composé d'une masse, fixée au milieu d'une tige raide nous ne tarderons pas à remarquer un certain défaut de correspondance entre ce système et les autres éléments mécaniques. En effet, en déplaçant avec la même vitesse

et dans la même direction les bouts (pôles) du ressort ou de l'amortiseur, nous n'éprouverons, évidemment, de la part de ces derniers aucune résistance. L'élément d'inertie de Lehr, au contraire, n'exercera pas de résistance dans les cas où le pôles (bouts de la tige) se déplaceront à vitesse égale en s e n s o p p o s é. C'est là ce qui nous empêche, en opérant avec l'élément d'inertie de Lehr, d'arriver aus règles simples pour l'obtention d'analogies (équivalences) élecrtomécaniques qui seront données ci-dessous.

A titre de développement de nos précédents travaux (²) nous proposons ici un nouveau système mécanique d'inertie, ne réagissant que sur les déplacements relatifs des pôles, système que nous appellerons élément d'inertie bipolaire ou, pour plus de brièveté, masse bipolaire (v. fig. 1).

Les tiges *ab* et *bc* sont réunies par des articulations entre elles (en le point *b*) et avec un certain point extérieur immobile (*a*). Au bout de la tige *bc*, en le point *c*, est fixée une masse *M*. Un des pôles de l'élément est fourni par le centre de la tige *ab*, l'autre par celui de la tige *bc*. Il va de soi qu'il ne faut considérer que les petites oscillations du système. Avec cet élément d'inertie le passage de n'importe quel schéma électrique à son équivalent (analogue) mécanique s'effectue sans difficulté.

Faisons observer qu'une masse unipolaire peut toujours être remplacée par une masse bipolaire dont le second pôle sera maintenu immobile. Le passage de la masse bipolaire à l'unipolaire n'est possible que dans le cas où l'un des pôles de la masse est rattaché à un point extérieur immobile.

3. Convenons d'admettre que le plan du dessin présente une résistance mécanique infiniment grande. On peut, par exemple, se figurer le plan du dessin sous forme d'une lame impondérable infiniment raide. Pratiquons dans cette lame des «fenêtres» en nombre correspondant à celui des éléments à réunir et plaçons-en chacun dans l'une de ces fenêtres, couplant l'un des pôles de chaque élément avec le bord supérieur de la «fenêtre» et l'autre avec le bord inférieur. Pour pouvoir réaliser le montage réciproque des éléments mécaniques exigible d'après le schéma, il est indispensable de sectionner la planche suivant un certain système de lignes. Ces lignes de sectionnement c.-à-d. les lignes de résistance mécanique égale à zéro, serviront de lignes de montage électrique, si nous passons du système mécanique a son équivalent électrique.

Supposons maintenant que nous ayons à opérer des couplages sur une lame conductrice. Tout comme auparavant, nous placerons les éléments du schéma électrique dans des fenêtres. Pour réaliser le schéma donné il faudra découper cette lame suivant un certain système de lignes. Il est évident que sur les lignes de sectionnement la résistance électrique sera égale à l'infini, c.-a-d, que les lignes de sectionnement seront des équivalents d'un couplage mécanique raide. Pour cette raison, le tracé de schéma électrique sera identique à celui d'un équivalent mécanique exécuté par le procédé ordinaire.

Le passage de l'ancien système de représentation au nouveau système ou inversement est fort simple, pour les couplages électriques aussi bien que mécaniques. Il est aisé de remarquer que le système des lignes de sectionnement est orthogonal au système des lignes de couplage, ces lignes réciproquement orthogonales n e v e n a n t à s e c o u p e r q u'a u m i l i e u d e s «f e n ê t r e s», à l'intérieur desquels sont placés les dits éléments.

L'établissement d'équivalences électromécaniques se ramène donc aux opérations suivantes:

A. Passage du schéma électrique à son équivalent mécanique. Chaque élément électrique est remplacé par un élément mécanique, les «pôles» subissant un déplacement par rotation de 90°. On construit un système de lignes orthogonales fermées et l'on considère ces lignes comme des lignes de couplage raide. L'un des «poles» de la force est à considérer comme en couplage raide avec un certain point extérieur immobile.

B. Passage du système mécanique à l'équivalent électrique. Les masses unipolaires sont remplacées par des masses bipolaires. Tous les points raidement fixés du système sont couplés les uns avec les autres.

Chaque élément mécanique est remplacé par un élément électrique, les pôles subissant un déplacement par rotation de 90°. On construit un système de lignes orthogonales fermées, lesquelles serviront de lignes de couplage électrique.

4. Les résultats des paragraphes précédents peuvent aisément être étendus aux mouvements rotatoires. De la masse unipolaire nous passons au moment d'inertie unipolaire.

En combinant le système de leviers de la fig. 1 avec le moment d'inertie unipolaire, nous passerons au moment d'inertie bipolaire.

Fig. 1

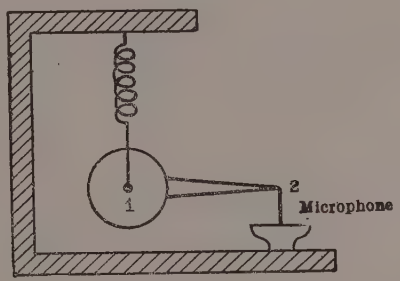

Fig. 2

En équilibrant une masse en rotation par rapport à son axe de rotation, nous obtiendrons des moments d'inertie «antiséismiques» unipolaires ou bipolaires, c.-a-d. des systèmes ne réagissant pas aux secousses venant heurter leur soutien.

A considérer un moment d'inertie antisésimique unipolaire dans lequel l'axe de rotation de la masse n'est pas fixé, nous voyons que ce système représente une combinaison d'une masse unipolaire avec un moment d'inertie bipolaire, le rôle de l'un des pôles revenant au centre de gravité de la masse (axe de rotation) et le rôle de l'autre pôle au bout de levier en couplage raide avec la masse.

Cet élément d'inertie complexe, jouissant de deux degrés de liberté, peut trouver son application en séismométrie.

En couplant à l'aide d'un ressort le centre de gravité (premier pôle) de l'élément d'inertie avec le soutien et le bout du levier (second pôle) avec un appareil enrégistreur (qui peut être en particulier un microphone), nous pouvons astreindre une seule et même masse à remplir deux fonctions différentes, à savoir: celle de masse inerte de séismographe (mouvements de trans-

lation) et de masse inerte d'un filtre mécanique de haute fréquence (mouvements de rotation).

Le schéma d'un appareil de ce genre que l'on pourrait appeler filtro-séismographe, est représenté en principe (fig. 2).

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PHYSIK

VERTEILUNG VON BEIMENGUNGEN BEI KRISTALLISATION

Von D. D. SARATOVKIN

(Vorgelegt von S. Wawilow, Mitglied der Akademie, d. 17. IX. 1935)

In allen sog. chemisch reinen Körpern, ist unbedingt ein gewisses Quantum von Beimengungen vorhanden, deren vollkommene Entfernung kein einziges von den üblichen Reinigungsverfahren erzielt.

In der vorliegenden Arbeit wird die Frage erörtert, wie während der Kristallisation aus einer Schmelze sich diejenigen Beimengungen verteilen, die keine Mischkristalle liefern, d. h. nicht in das Gitter eines wachsenden Kristalls eintreten, doch ebensowenig den mechanischen Beimengungen zugezählt werden können, d. h. nicht als feine Fremdkörper in der Schmelze schwimmen. Alle, über diese Frage in der Literatur vorhandenen Angaben können auf folgende Weise zusammengefasst werden: «Kühlt die Schmelze langsam genug an dem einen Ende ab, so weisen die Beimengungen die Tendenz auf, sich in dem Teile anzusammeln, der zuletzt kristallisiert. Wenn die Kristallisation zu schnell vor sich geht, so dass die Körner der Kristalle sich über die ganze Flüssigkeit hin bilden, so müssen sich die Beimengungen zwischen den Kristallen ansammeln. Geht die Kristallisation noch schneller vor, so müssen sich einige Beimengungen an den Grenzen der Mosaik ansammeln, aus welcher jeder mikroskopische Kristall besteht»(1).

Die Verdrängung der Beimengungen durch den wachsenden Kristall führt sich auf Diffusion der Beimengungen von der Seitenfläche zurück. Es sei dm die Menge des diffundierenden Stoffes, welche während der Zeitspanne $d\tau$ die Oberfläche S senkrecht zum Konzentrationsgradienten dc/dx durchströmt; dann ist

$$\frac{dm}{d\tau} = -DS\frac{dc}{dx},$$

wo D der Diffusionskoeffizient ist. Das Minuszeichen weist darauf hin, dass $dm > 0$ ist, wenn c mit zunehmendem x abnimmt. Der positive Gradient dc/dx bedeutet aber Zunahme der Konzentration beim Anwachsen der Seitenfläche eines Kristalls, was nur in dem Falle eintreten kann, wenn die Menge der ausscheidenden Beimengungen kleiner ist, als die der auf Kosten der Verdrängung durch die Seitenfläche des Kristalls hinzukommenden. Mit anderen Worten, kann die Diffusion der Beimengungen nur bei Vorhandensein eines Konzentrationsgradienten erfolgen; ein solcher entsteht aber nur, wenn die Geschwindigkeit des Anwachsens der Seiten-

fläche grösser, als die der von letzterer diffundierenden Teilchen wird. Auf diese Weise befindet sich vor der wachsenden Seitenfläche des Kristalls immer eine Schicht mit gesteigerter Konzentration der Beimengungen.

Das Wachsen der Seitenfläche wird durch die aus der Schmelze sich darauf ablagernden Molekeln bewirkt.*. Das Vorhandensein einer Schicht mit gesteigerter Konzentration erschwert dieses Anwachsen.

Betrachten wir zwei anstossende Seitenflächen eines Kristalls AB und BC (Abb. 1); die Geschwindigkeiten ihres Wachsens v_1 und v_2 können, im allgemeinen, verschieden sein. Nach einer gewissen Zeitspanne werden sie die Lage $A'O$ und OC' einnehmen. In der Umgebung des Punktes C' werden sich alle Beimengungen ansammeln, die sich anfänglich auf dem Abschnitt CC' befanden; ebensoviel Beimengungen wird es beim Punkte B' geben, da $CC' = BB'$ ist. Beim Punkte D' werden sich aber weniger Beimengungen vorfinden, da $DD' < BB'$ ist, und die Beimengungen, die sich auf dem Abschnitte DD' befanden, von dem Augenblick an, wo der Kristallrand im Punkte D erscheint, verdrängt zu werden beginnen. So wird denn auf dem Gebiete der Seitenfläche OB' die Konzentration der angesammelten Beimengungen geringer sein, als auf der ganzen übrigen Seitenfläche, und zwar umso gerin-

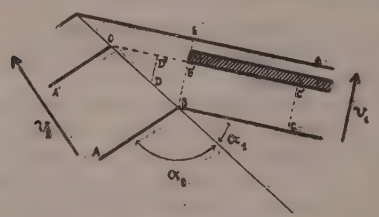

Fig. 1

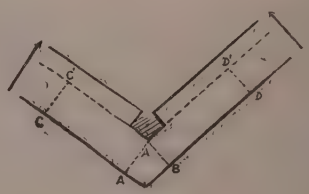

Fig. 2

ger, je näher dem Rande. Das Gebiet OB' wird geringere Hindernisse für sein Wachsen antreffen, als die übrige Seitenfläche, es überholt dieselbe und, wenn der daraus entstehende Vorsprung $B'E$ grösser als die Schichtdicke der sich an den Seitenflächen angesammelten Beimischungen wird, erweitert er sich seitwärts, indem er zwischen sich und dem Kristall eine Schicht von Schmelze mit gesteigerter Konzentration der Beimengungen einpresst. Weiterhin wird sich der Prozess wiederholen. Dasselbe wird auch an einer beliebigen anderen Seitenfläche abspielen. Im Ergebnis erhalten wir einen von jeglichen Beimengungen freien «Stamm» mit «Zweigen», die unter sich Zwischenschichten mit Beimengungen einschliessen Die vom «Stamme» und seinen «Zweigen» gebildeten Winkel stehen mit den Wachstumsgeschwindigkeiten der Seitenflächen in der Beziehung

$$\frac{\sin \alpha_1}{\sin \alpha_2} = \frac{v_1}{v_2}.$$

Betrachten wir den Fall, wo in einem grösseren Raume der von der einen Seite durch die Schmelze begrenzte Kristall sich befindet. Fällt die Grenze zwischen dem Kristall und der Schmelze wegen der thermischen Bedingungen mit keiner der natürlichen Seitenflächen des wachsenden Kristalls zusammen, so muss diese Grenze, beim Anwachsen des Kristalls, eine Oberfläche mit den von den natürlichen Seitenflächen begrenzten Vorsprüngen dar-

* Hier, wie auch fernerhin, versteht man unter der Molekel den allgemeinen Fall. Im Einzelfalle können es Atome oder Ionen sein.

stellen. Bei näherer Betrachtung zweier benachbarter Seitenflächen, die unter sich einen inneren Winkel bilden (Abb. 2), bemerken wir, dass im Punkte A' mehr Beimengungen vorhanden sein müssen, als im Punkte C' oder D', nämlich letztere müssen an Zahl der Summe der ersteren gleich sein. Im Raumwinkel selber werden wir eine besonders grosse Steigerung der Konzentration erhalten. Der Wachstum der Seitenflächen wird in der Nähe des Winkels erschwert sein. Es bilden sich Vorsprünge, wie auf der Abbildung gezeigt ist, und die Stelle mit der höheren Konzentration wird als ein inselartiger Einschluss erscheinen. Auch weiterhin wird sich dieser Prozess in derselben Weise wiederholen. So wird sich denn für den Kristall folgendes Aussehen ergeben (Abb. 3). Ein jeglicher Aussenrand zwischen zwei Seitenflächen wird auf seinem Wege eine Spur ohne Beimengungen zurücklassen; jeglicher Innenrand dagegen — eine sehr verunreinigte Spur. Dazwischen, in den zu den Seitenflächen parallelen Richtungen, werden feine Zwischenschichten mit Beimengungen vorkommen. Auf Grund unserer Ausführungen müssen diese Zwischenschichten (feine Schraffierung) immer bestimmten kristallographischen Richtungen entsprechen, die Linien der grössten Reinheit und der maximalen Verunreinigung (grobe Schraffierung), dagegen, den Richtungen der Randverschiebungen. Sollte den Abkühlungsbedingungen gemäss, das Verhältnis der Wachstumsgeschwindigkeiten der betreffenden Seitenflächen veränderlich sein, so führt das zu einer Krümmung der Linien von grober Schraffierung.

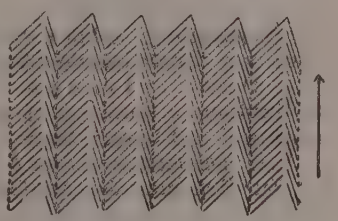

Fig. 3

Die Grösse der Zähne auf Abb. 3 hängt vom Temperaturgradienten ab. Je kleiner dieser Gradient, um so grösser die Entfernung zwischen den Winkeln; folglich, je langsamer die Abkühlung zustande kommt, desto grösser die Entfernung zwischen den Linien grober Schraffierung. Die feine Schraffierung (Zwischenschichten) ist ebenfalls bei langsamer Abkühlung besser zu sehen.

Betrachten wir nun die Frage, was in dem Teile der Schmelze vorgeht, welcher einen erhöhten Gehalt an Beimengungen bewahrend, von kristallinischer Masse eingeschlossen wird. Mit Steigerung des Konzentrationsgrades der Beimengungen, sinkt die Kristallisationstemperatur der Grundkomponente sehr erheblich, weshalb die Zwischenschicht, die von allen Seiten von Kristallsubstanz umgeben ist, noch flüssig bleibt. Mit der Herabsetzung der Temperatur des Kristalls, setzen sich die Molekeln der Grundkomponente aus dieser Zwischenschicht auf die Kristallsubstanz ab. Die Konzentration der Beimengungen steigt. Sobald die Zusammensetzung der Zwischenschicht der Eutektik von Beimengung mit Grundsubstanz entspricht, erstarrt die ganze Zwischenschicht. Also bestehen die Zwischenschichten im Kristall aus Eutektik. Dieser Umstand erhöht recht stark das spezifische Gewicht der Zwischenschichten gegenüber demjenigen der Beimengungen. So z. B. verleiht eine Beimengung von 0.013% Antimon in Blei einem Zehnte des Bleikristalls den Charakter von Zwischenschichten der Eutektik Sb — Pb.

Auf Grund der angeführten Erwägungen lässt sich eine Reihe von allgemein bekannten Erfahrungstatsachen erklären.

1. Lockerheit langsam schmelzender Kristalle. Eine derartige Erscheinung kann man auf Flusseis, welches im Frühling langsam schmilzt, beobachten, wenn es locker wird und in zur Eis-

schollenfläche senkrecht gelegene Stäbchen zerfällt. Die Achsen dieser Stäbchen schmelzen zuletzt. Zuerst schmelzen die Zwischenschichten, welche die im Wasser aufgelösten Mineralbeimengungen enthalten.

2. Widmannstettensche Struktur. Diese Struktur ist dadurch charakterisiert, dass in den Ferritkörnern, parallel den Würfel-, Oktaeder- und rhombischen Dodekaederflächen, Zwischenflächen von Perlit zu liegen kommen, d. i. von Eutektik des Ferrits mit Zementit. Strukturen von solchem Typus trifft man nicht nur im System Eisen—Kohlenstoff, sondern auch in anderen Legierungen.

3. Ätzfiguren. Bei Einwirkung von Säuren auf die Oberfläche von metallischen Kristallen bildet sich auf der Oberfläche ein Netz von Mustern, die der kristallographischen Struktur des betreffenden Kristalls entsprechen. Die Erscheinung erklärt sich durch die Verschiedenheit der Lösbarkeitsgrade von Zwischenschichten und Kristall.

4. Dendritische Kristallisation. Alle Arten von Kristallbildung in Form von tannenartigen oder sonstigen baumartigen Mustern lassen sich auf Grund des Schema von Abb. 1—3 erklären.

Verf. beobachtete unter dem Mikroskop die Kristallisation von Hyposulfit. Zur Regelung der Abkühlungsgeschwindigkeit wurde unter dem Objektivglas ein durch den Strom erhitzbarer Draht angebracht. Bei Verunreinigung des Hyposulfits durch Chlorkobalt ($CoCl_2$), ergeben sich Verbindungen, die zwar keine Mischkristalle bilden, dafür aber eine scharf ausgesprochene Färbung aufweisen; bei Herstellung einer entsprechenden Abkühlungsgeschwindigkeit gelang es uns die Steigerung der Konzentration der Beimengungen vor einer wachsenden Seitenfläche sowie die Besetzung der Zwischenschichten zu beobachten (Abb. 1—3).

Im Falle von Überfluss an Wasser wurde letzteres auch zu einer Beimengung; ein vollkommenes Erstarren der Einschlüsse mit gesteigerter Konzentration des Wassers fand bei Zimmertemperatur nicht statt; daher waren die eingeschlossenen «Inseln» (Abb. 3) sehr gross, regelmässig begrenzt und kettenartig aneinandergereiht.

Die Beobachtungen wurden mit einem Polarisationsmikroskop ausgeführt.

Zum Schluss spreche ich meinen Dank für eine Reihe wertvoller Ratschläge den Professoren Dr. V. Kuznecov, A. Dobrovidov, P. Tartakovskij, M. Usanovič und den Dozenten: M. Bolšanina, J. Grdina und I. Molodych aus.

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TECHNICAL PHYSICS

DRYING AND STERILIZING WOOD AND OTHER STRUCTURAL MATERIALS WITH INFRA-RED RAYS

By P. N. ULJANOV

(Communicated by G. A. Nadson, Member of the Academy, 9. X. 1935)

In Switzerland, an attempt has been made to dry structural wood with electricity. As may be seen from technical publications (1), the attempt gave good results and proved the method to be commercially sound. In the USSR, where building operations are carried on all the year round and raw timber has to be used in erecting wooden structures, the engineer is faced with a still more complex problem, viz., the elaboration of new dynamic methods for sterilizing destructive fungi, drying wooden structures and materials.

One of the physical peculiarities of infra-red rays is that they are but slightly absorbed by air and very considerably absorbed by structural materials. 1% of infra-red rays, $\lambda=0.6—3\ \mu$ is absorbed by a layer of damp air of 910—980 m (2). The measurements made by Zelenin and myself of the absorption coefficents of the continuous spectrum of intra-red radiation with a maximum of energy at $\lambda=1.8\ \mu$ are given in the table below.

Object exposed to radiation	*Belatus edulis fungus*	Pinewood	Stucco	White brick	Red brick	Felt	Tarred paper	Newspapers
Percentage of energy reflected at an angle of 45°	—	4.1	5.26	8.5	6.66	4.17	2.85	6.66
Percentage of energy that penetrated through a layer 10 mm thick	2.53	0.4	2.85	3.3	2.75	2.6	2.9	1.0
Percentage of energy absorbed by a layer 10 mm thick . .	*ca.* 93.0	95.5	91.89	88.2	90.59	93.23	94.25	92.34

The results obtained for reflected energy show a good agreement with the data of Barker (3), who estimates at 5—10 the percentage of infra-red rays reflected by walls, floors and ceilings. Ashkinasi has estimated the penetrativity of infra-red rays $\lambda=1.2—0.9\ \mu$ for water $D=1$mm at 2.2—38.7%, which gives an order of magnitude agreeing with ours. On the other hand, a

Depth of penetration	1	2.6	16	23	39	50	60	mm
Observed penetration of energy.	22	15	11	9	4	—	—	$\times 10^3 \frac{\text{erg}}{\text{cm}^2 \text{ sec}}$
Calculated penetration of energy	21.7	19.8	11.1	8.3	4.0	2.5	1.6	$\times 10^3 \frac{\text{erg}}{\text{cm}^2 \text{ sec}}$

computation of the data obtained by us for the absorbability of infra-red rays in absolute units gives results similar to those obtained by Weigert (5) in his observations of water. The latter has proved that the continuous infra-red spectrum is only absorbed by a layer of water 5 or more cm thick.

Fig. 1 shows the permeability of wood to infra-red rays. The following table shows the power of second infra-red radiation from heated pin

Temperature of wood surface, °C	32	40	55	75	95	110
Power of radiation, $\frac{\text{erg}}{\text{cm}^2 \text{ sec}}$	6.3	25	50	70	87	104

wood, measured by Zelenin. It should be taken into account that the depth of penetration of active infra-red rays depends on the power of the radiation supplied. The power, and, consequently, the depth of penetration, may be increased in the case of brick and stucco. Wood, however, sets a limit in this respect, since there exists a permissible maximum of heating for the object exposed to radiation. The power of infra-red rays used in the measurements given above was such as not to raise above 70—65° the temperature of that side of the object that was exposed to radiation.

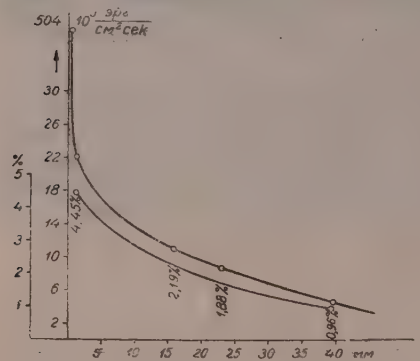

Fig. 1. Permeability of pinewood (in mm of wood). $E = 0.045\ E_0 l^{-0.044 d}\ \frac{\text{erg}}{\text{cm}^2 \text{ sec}}$. . . Infra-red spectrum with maximum energy $\lambda = 1.8\ \mu$.

Both the permeability and the absorption of infra-red rays are selective, differing as they do according to the wave lengths of the spectrum. For cellulose, the maximum permeability is in the region of the spectrum $\lambda = 1.0—2.5$ and $4.0—5.5\ \mu$ (4). As established by us, the maximum of permeability for wood is $\lambda = 2.7 — 3.0\ \mu$.

In drying wood and other structural materials, it is essential to follow up the thermal effect of infra-red radiation.

In any given layer of wood, the rise of temperature will follow the equation: $T = aE$, where T is the temperature, E is the energy, a is the compound coefficient of proportionality, due to absorption, to the thermal capacity of wood and the choice of units of measurement.

Given the permeability shown in Fig. 1, the heating of wood due to absorbed radiant energy gives, after 30 min. of radiation a temperature of 138° C at a depth of 1 cm, 1°C at a depth of 3 cm, and 0.25°C at a depth of 6 cm.

The difference of temperature thus set up in the thickness of the wood will give rise, in accordance with the well-known equation of thermal conductivity, the heating of deeper layers by contact. For the same reason, we shall have, within the same half hour, a shifting of temperature towards the centre of the wood, the values being 93° at a depth of 1 cm, 12° at a depth of 3 cm, and 2.2° at a depth of 6 cm.

With the temperature of ambient air of 18°, the cooling of the more heated portion of the wood (including radiation) gives, within the same period of time, an ultimate temperature of only 47° above the temperature of the air for a depth of 1 cm, 9° ditto for a depth of 3 cm, and 2° ditto for a depth of 6 cm.

With a judicious choice of the infra-red spectrum, a more powerful radiation and a longer period of radiation, the temperature within the wood may be more than doubled or trebled as compared with convection heat (*cf.* Fig. 2 and 3).

In this case, the difference from heating by convection is as follows: 1) contact heat will become superimposed on the heat generated in the central and outer layers due to absorbed radiant energy, and 2) the temperature of the ambient air is always lower than that of wood and will therefore shift the maximum of temperature within the wood towards the innermost layers.

In wood drying, the thermal gradient is as high as 40—20° per 1 cm in the outer layers and 20—8° in the innermost layers, and this reduces the drying time to a few hours.

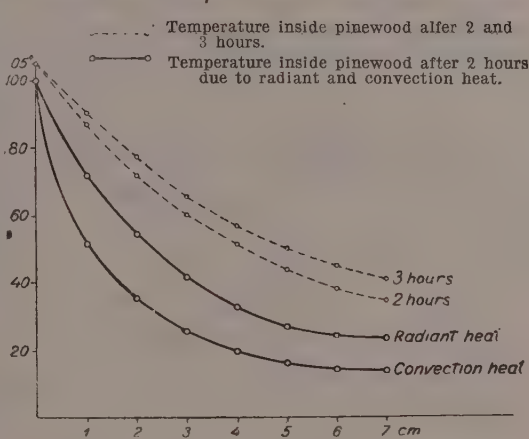

Fig. 2. Symbols used: maximum energy at $\lambda = 1.9\ \mu$.

The gradient of humidity is as high as 5—6% per 1 cm.

If, in drying, the heating of the outer layers is gradual, from 40 to 100° within 30—40 min, so as to bring into action not only radiant, but also contact heat, no visible deformation of the wood is noticeable in drying. The observed absence of overstrains in the wood is due not only to an accelerated diffusion of humidity from the wetter into the drier layers, but also, apparently, to the pressure of saturated water vapour in the wet layers of wood and the simultaneous pumping action of non-saturated vapour in the dry layers. An essential part is undoubtedly played in the drying of wood by the steam film which forms round the wood owing to the cooling in the air of the vapours issuing from the wood (see Fig. 4).

The drying of wood is in itself one of the most effective methods of destroying wood fungi.

In drying with infra-red rays at a depth of 7—6 cm, lethal temperatures

can also be obtained for fungi, say, by exposure to radiation of one side only 52°. The thermal effect of infra-red rays is only a secondary one.

Starting from the electro-magnetic nature of infra-red rays it may be assumed that their primary effect is electro-magnetic. The writer had an opportunity of observing this in a series of experiments made by him conjointly with A. Goldfeld, when they exposed to radiation the eggs of the frog (*Rana temporaria*). The radiation was produced with infra-red rays with a maximum of energy at $\lambda = 2$ μ; the temperature of the object was below 24° and was maintained by air cooling. Subsequent growth takes place at a temperature of 18°. Temperature control eliminated the thermal effect.

The experiments proved that, with an increasing power of radiation, the development of tadpoles from the eggs took place in close accordance with the law of Arnt-Schultz-Nadson (6), passing through the stages of activation, inhibition and lethal action. The biological action of infra-red rays follows that law.

A number of references to both the inhibitory and the lethal effect of infra-red rays are to be found in literature. The inhibitory effect of infra-red rays bas heen noticed by physio-therapeutists, who use these rays in the treatment of chronic inflammatory processes (7). The lethal effect of infra-red rays is confirmed by the experiments of Wiesner (8), who investigated

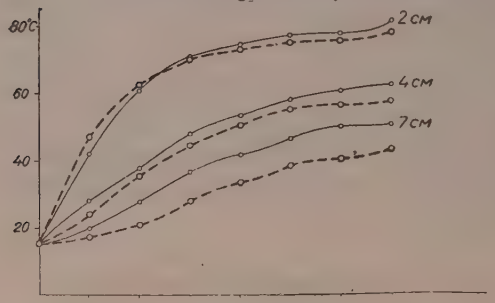

Fig. 3. Temperature inside pinewood (absolutely dry) when exposed to infra-red radiarion. (Temperature of outside surface 105°).

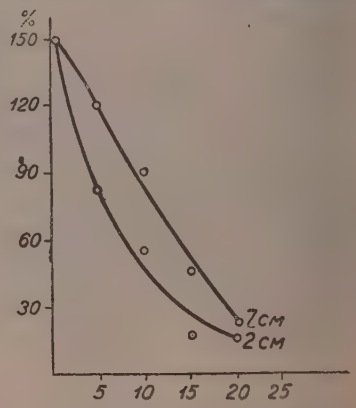

Fig. 4. Moisture in drying 15 cm pinewood with infra-red spectrum with maximum energy $\lambda = 1.9$ μ, with two opposite surfaces exposed to radiation.

their bactericide action. In Wiesner's experiments, the effect of infra-red rays proved more potent than the thermal effect or than the ultra-violet ray effect.

A direct investigation of the fungicide action of infra-red rays, carried out by E. Nazarova on my suggestion and in which I took part, has fully confirmed the theoretical data set forth above and is in complete agreement with Wiesner's observations.

The total duration of exposure to radiation was 24 hours; it was divided into four periods of 6 hours each, separated by intervals of 18 hours. A stream of infra-red radiation with the maximum energy at $\lambda = 2.7$ μ of a power of $5.0 \cdot 10^5$ erg/(cm^2 sec) retarded the growth of fungi by 2—3 months.

A power of 9.0—11.0·10⁷ erg/(cm² sec) achieved a complete sterilization of the fungi used in the experiment, viz., *Merulius Lacrymans* Schum., *Poria vaporaria* Pers. and *Coniophora cerebella* Schr.

With the same power and period of exposure, infra-red spectra having their maximum energy at $\lambda=1.8\ \mu$ gave a lethal effect of only 75%, and those having their maximum energy at $\lambda=4.3\ \mu$ of only 30%.

The action of heat in the thermostat, maintained at a temperature of 93—105°, during the same period of time, proved lethal to the extent of 40—50%. The wood placed in the thermostat had all its surfaces exposed to the action of convection heat.

When only one side was exposed to radiation, we had the temperature of 92—100° only at the side so exposed, the temperature at the opposite side being only 27—47°.

It is obvious that during exposure to radiation the heat only increased the electro-magnetic biological action of the infra-red rays. This is particularly emphasized by the selective action of the sectors.

It might be of interest to cite some facts relating to the use of infra-red rays in practice.

In drying structural materials used in building, the estimate of the requisite amount of energy should be based on the thermal effect of infra-red rays. Such an estimate made by the writer together with Eng. Alferyev, for the drying of the floor of a standard classroom last summer, gave a total of 740 kilowatt hours. The estimate is based on the actual moisture of the floor wood of 67%, the weight of water to be removed being 924 kg. With an expenditure of power equal to 35 kilowatt hours, 20—24 hours would suffice to have such a floor dried with infra-red rays.

As regards sterilization, it might be of interest to give the temperatures observed by Zelenin within the floor during exposure to radiation, viz.:

Surface of the stucco exposed to radiation 105—110°
Front surfaces of the wood protected by laths and mat 75—80°
Temperature in the wood:
at a depth of 4.0 cm from surface exposed to radiation 65°
» 5.5 » » » » 52°
» 7.0 » » » » 50°
» 8.0 » » » » 47°

Here is a passage from the opinion of the superintendent of building operations: «This method of drying eliminated the delays in completing the floor, which are unavoidable with the natural method of drying. It reduced the drying to a few hours and rendered it safe and easy of execution».

In the winter of 1934, the drying of plaster in the Biogroup Building of the Academy of Sciences, carried out by K. Uljanov by way of trial at an air temperature of 3—50 and with 60% moisture in the air, made it possible to dry a laboratory room in 24 hours to such an extent as to admit the beginning of finishing and painting work the very next day after drying. A 5 kW stove was used. Attempts of drying with coal-fed stoves caused the plaster to absorb moisture and partly to come off.

The attempted drying and sterilization of metre lengths of wooden beams thoroughly wetted and affected by fungi gave quite satisfactory results only down to a depth of 7—8 cm. The drying of the central zone of a panel of 30×40 cm was much slower and was mainly attributable to the fissures forming between the blocks of the panel.

As a result of the work done, the physical and technical requirements to be filled by infra-red generators may be summarised as follows. In order to ensure first of all the maximum penetration of the rays into the thickness of the wood, we brought the temperature of the heating of the radiator up to 800°. This temperature is also high enough to ensure a profitable yield of radiant energy. In irradiating plane surfaces, reflectors of the footlight type were used, whereas searchlight type reflectors were given the preference in irradiating corners.

Conclusions

The simplicity of design of the generators, as well as the low cost of drying and sterilizing structural materials with infra-red rays, is a guarantee of the wide-spread success of this method.

The high penetrativity of infra-red rays and their absorption by building materials make it possible to effect the drying and sterilizing in any weather conditions.

The action of infra-red rays has been tested in solving a number of technical problems connected with the drying and sterilization of erected structures, and has been found to give satisfactory results.

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ENGINEERING

ON THE STABILITY OF EARTH SLOPES SUBJECTED TO THE ACTION OF GROUND WATER FLOW

By R. R. ČUGAIEV

(Communicated by N. N. Pavlovskij, Member of the Academy, 4. XI. 1935)

The present communication sets forth the results of the writer's investigation into the question of the stability of hydrotechnical earth structures. Such a structure is characterized by the fact that the earth it is made of is usually saturated with either still or moving water, which gives rise, in a general case, to the following a d d i t i o n a l forces acting on the framework of the soil, viz.: the volume forces of water uplift and percolation, as well as the surface forces of capillary pressure. The presence of these additional forces complicates to a great extent the theoretical analysis of the stability of the soil masses of which the slopes of the structure are made, the computation methods devised by a number of authors [Sven Hulten, Krey, Fellenius, Terzaghi (1), and others] for dry earth being in most cases inapplicable to hydrotechnical structures. It is to be noted that, in spite of the great interest this problem has assumed in connection with recent technical developments, it has scarcely been dealt with in literature.

In discussing below some essential problems of the computation of hydrotechnical earth structures, we shall adopt Petterson's suggestion, according to which the surface of rupture of the slope may be assumed, for purposes of computation, to have a circular cylindrical shape. In doing so and in formulating the problem as in the case of «dry» slopes, we shall assume that the degree of stability of the slope under consideration may be appraised by the safety factor for stability (β) of a certain body of ground bounded up by «t h e m o s t d a n g e r o u s» cylindrical surface of rupture, in the assumption that such a «most dangerous surface», characterized by the minimum value of the safety factor for stability, ought to be determined by successive approximations. The problem being thus formulated, we shall certainly find it sufficient to discuss below the stability of the slope only in respect of o n e arbitrarily chosen circular surface of rupture (*cf.* below circumferences *BOF* and *FOE* in Fig. 1 and 2 respectively).

It may be observed that we consider below a «plane problem» only, no account being taken of the elasticity of the ground framework, which sometimes gives rise to a special unsettled flow of the water (2).

§ 1. Let us consider an earth slope (Fig. 1), and, disregarding the capillary uplift (the ground is assumed to be sufficiently coarse-grained), let us

suppose that the water in the reservoir rapidly (practically instantaneously sdescends from position (*1*) to position (*2*). Such a fall of the ground water level would evidently give rise to an unsettled percolation of water, the initial moment of which (*cf.* stream lines shown by arrows on Fig. 1) would be the most dangerous one for the stability of the slope. The plotting of lines of equal heads (h) corresponding to the initial moment of percolation may be done by the well-known method of electrohydrodynamic analogies (the *EHDA* method) suggested by Academician N. N. Pavlovskij (3). According to that method, the following potentials of electricity have to be noted on the staniol plate bounded up by the line *ABDN* (taken on any scale), viz.: along the lines *DN* and *ABC*, respectively, certain potentials W_1 and W_2 ($W_2 < W_1$), and along the line *DC*, the linear change of potential from the value W_1 (at point *D*) to the value W_2 (at point *C*). After finding on the *ENDA* apparatus, in the boundary conditions stated, the system of h lines that characterize the ground flow under consideration, we have to determine the field of force obtaining within that body of the ground the stability of which si being considered; for this purpose we examine the lower portion of that body (*DMOB*) immediately subjected to the action of the water. Resorting with that end in view to the theory of percolation forces of Academician N.N. Pavlovskij*, we follow that author's suggestion as to the plotting, from the available h lines, of equipotential lines (Φ) of the field of the «resultant» forces (4), where the «resultant» forces mean the resultants of the gravity and ground forces, of the uplifting forces of water and of the percolation forces. The system of Φ lines so obtained (Fig. 1) is found to be, as shown by computations, very close to a system of straight parallel and equidistant lines. This makes it easy te deterime, from the Φ lines, with a precision fully adequate for practical purposes, both the average specific value of the «resultant» forces γ_0 and their average direction (*cf.* angle θ, Fig. 1). By assuming for a portion of the sliding body *DMOB* the corresponding values of γ_0 and θ, we thus reduce our calculation, as this may be seen, to the consideration of a case where the body in question is subjected to the action of only two systems of forces acting in different directions, viz.: a vertical system of gravity forces (in the *DFEM* portion), and an inclined system of «resultant» forces (in the *DMOB* por-

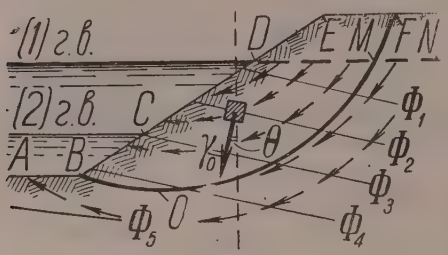

Fig. 1.

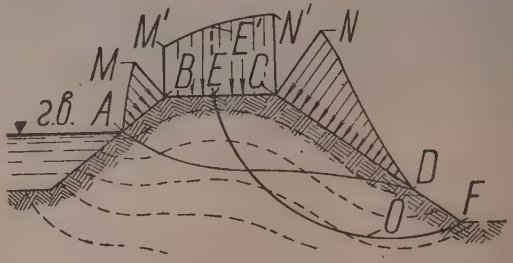

Fig. 2.

* The main points of the work of Academician N. N. Pavlovskij referred to have now been embodied into a report submitted by him to the XVI International Navigation Congress at Brussels in 1935.

tion). The required value of the safety factor β may be computed from the formula (4)*:

$$\beta = \frac{f[\gamma I_1' + \gamma_0' I_2' - 2\gamma_0'' I_2'' + \gamma_0''' I_2'''] + \frac{c_1}{b_1} S_1 + \frac{c_2}{b_1} S_2}{\gamma I_1'' + 2\gamma_0'' I_2 + \gamma_0 I_2' - \gamma_0'' I_2'''}.$$

§ 2. We shall now examine the case of the downstream slope of an earth dam and shall assume the soil in question to be characterized by a fairly high capillary uplift. Under these conditions, the menisucs surface will be seen to be established along the line *ABCD* (Fig. 2), and, since the «head» concept is fully applicable to the capillary zone, it is evident that the whole profile of the dam up to its top (*cf.* stream lines shown with dotted lines in Fig. 2) is saturated with flowing water. As may be seen, we have here a special case of water motion differing in nature both from the «non-head» and the «head» movement of water. It is easy to plot the above ground flow by using the *EHDA* method, and, after the *h* lines have been found, it is easy to calculate the line of atmospheric pressure *AD*, above which the capillary zone lies. Furthermore, the *h* lines being known, it is easy to make a diagram of capillary pressures obtaining within the sliding body *CFOE* on the surface of the meniscuses (*cf.* areas *EE'N'C* and *CND*, Fig. 2). It is equally easy to find, as explained in the preceding section, the values of γ_0 and θ, which determine the intensity and the direction of the «resulting» forces obtaining within the whole of the sliding body *CFOE*. Upon determining in this manner all the necessary forces and in order to simplify the static computation, we divide the diagram of capillary pressure *EE'N'C* into two component diagrams in such a manner as to obtain only two systems of forces acting on the sliding body in different directions. Thereafter, the factor β can evidently be easily calculated from the above formula.

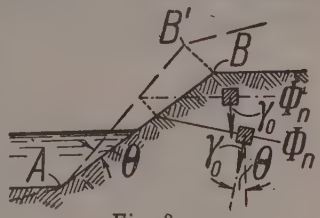

Fig. 3.

§ 3. Earth slopes are sometimes known to collapse after rainfall. The causes giving rise to such collapses may in some cases be as follows (*cf.* Fig. 2). Rain water collecting on surface *BC* and also flowing along surfaces *AB* and *CD*, destroys, so to say, the meniscus surface, so that such ground water as lies above the line *AD* loses its capillarity and, impelled by gravity forces, begins to flow downwards, appearing on the surface of the slopes. In some cases (in the absence of percolation coming from the water reservoir), such a flow of water may be regarded as unsettled percolation, exactly similar in appearance to that represented in Fig. 1. Therefore, in testing the stability of a slope «against rain», the values of γ_0 and θ are determined as explained in § 1; as to the static calculation, it may be done in the following manner: the slope under consideration (Fig. 3) is rotated about point *A* through the angle θ, *i. e.* in such a way that the Φ lines should assume a horizontal position, while the volume forces γ_0 should act in a vertical direction. As a result of such rotation, we obtain, instead of the actually given slope *AB*, an imaginary slope *AB'*, characterized by a slightly greater height and by a somewhat greater steepness. The imaginary slope *AB'* subjected to the action of volume forces γ_0, is evidently quite equi-

* For an explanation of the notations *cf.* ibid.

stable to the given slope AB, which is subjected to unsettled percolation due to rainfall. We may therefore use in our calculation, instead of the real slope AB, the imaginary slope AB', it being possible to make that calculation in the usual manner, by methods applicable to «dry» slopes ([1],[2]), but assuming the volume weight of the soil te be equal to γ_0.

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CHEMIE

DIE THERMISCHE ZERSETZUNG DER DIMETHYLCYKLOHEXANE

Von J. A. ARBUZOV und B. M. MICHAJLOV

(Vorgelegt von N. D. Zelinskij, Mitglied der Akademie, d. 17. IX. 1935)

In der Abhandlung von N. D. Zelinskij, B. M. Michajlov und J. A. Arbuzov (1) wurde über die Resultate einiger thermischen Zersetzungsversuche mit Cyklohexan und seinen einfachsten substituierten Homologen—Methylcyklohexan und Äthylcyklohexan—in überschüssigem Wasserdampf berichtet.

Diese Zersetzungsmethode von Kohlenwasserstoffen war aus folgenden Gründen angenommen worden. Bei der Verdünnung des zu zersetzenden Stoffes mit Wasserdampf wird eine kurze Kontaktierungszeit, sowie auch eine bedeutende Herabsetzung des Partialdruckes des zu zersetzenden Stoffes erreicht. Unter solchen Bedingungen wird keine Kohlenabscheidung an den Wandungen des Reaktionsgefässes und auch keine nennenswerte Bildung teerartiger Kondensationsprodukte beobachtet. Die flüssigen Zersetzungsprodukte bestehen selbst bei hochprozentigem Zerfall (40—60%) hauptsächlich aus unverändertem Kohlenwasserstoff. Butadien, das zweifellos ein primäres Zerfallprodukt des Cyklohexans und seiner Homologen darstellt, verfügt über eine geringe thermische Widerstandsfähigkeit und tritt unter gewöhnlichen Bedingungen der Pyrolyse leicht in Polymerisations- und Kondensationsreaktionen ein. Bei der Zersetzung von Cyklohexan in überschüssigem Wasserdampf enthielten die gasförmigen Zersetzungsprodukte eine bedeutende Menge Butadien (15—19%), während bei der Zersetzung von Cyklohexan ohne Verdünnugsmittel, die von B. A. Kazanskij und A. F. Plate (2) in ihren Versuchen durchgeführt wurde, der Butadiengehalt in den gasförmigen Zersetzungsprodukten nur 6—6.4% beträgt.

In der vorliegenden Abhandlung werden die Resultate der thermischen Zersetzung doppeltsubstituierter Homologe des Cyklohexans, nämlich des 1,2-, 1,3- und 1,4-Dimethylcyklohexans angeführt, deren nichtkatalythische Zersetzung bisher noch keiner Untersuchung unterzogen wurde.

Die Zersetzung wurde, ebenso wie bei der vorhergehenden Arbeit, in überschüssigem Wasserdampf bei Temperaturen von 675—750° vorgenommen. Die erhaltenen gasförmigen Zersetzungsprodukte zeigten in allen Versuchen einen recht beständigen Prozentsatz ungesättigter Kohlenwasserstoffe (47—52%), Wasserstoff (15—18%) und gesättigter Kohlenwasserstoffe (30—36%). In Bezug auf den Gehalt an Äthylen, Propylen und Butylenen zeigen die bei der Zersetzung verschiedener Isomere erhaltenen

Gase merkliche Differenzen. Während in den Versuchen mit 1,2-Dimethylcyklohexan der Äthylengehalt des Gases (23—27%) die Summe des Gehaltes an Propylen und Butylenen (13—14%) bedeutend übertrifft, ist er in den Versuchen mit 1,3-Dimethylcyklohexan ungefähr ebenso gross (19—22% bzw. 20—22%), und in den Versuchen mit 1,4-Dimethylcyklohexan sogar etwas kleiner als die Summe des Gehaltes an Propylen und Butylenen (17—19% bzw. 19—24%). Aus allen 3 isomeren Dimethylcyklohexanen erhält man Butadien, aber in bedeutend geringeren Mengen als aus Cyklohexan und seinen monosubstituierten Homologen. Bei der Zersetzung von 1,2-Dimethylcyklohexan bei 725° wurde Butadien mit einer Ausbeute von 20.7 Mol % (auf die zersetzte Substanz bezogen) erhalten, im Versuch mit 1,3-Dimethylcyklohexan bei 700° wurde 18.4 Mol % Butadien gewonnen und aus 1,4-Dimethylcyklohexan bei 750°—23.9 Mol % Butadien. Bei der Zersetzung eines Gemisches isomerer Dimethylcyklohexane, das durch Hydrierung des im Handel befindlichen Xylols erhalten war, wurde Butadien mit einer Ausbeute von 22.5 Mol % Butadien gewonnen. Durch fraktionierte Destillation der in den Versuchen mit 1,2- bzw. 1,3-Dimethylcyklohexan erhaltenen flüssigen Zersetzungsprodukte wurden niedrigsiedende Fraktionen (30—50°) isoliert. Durch Analyse im Korotkov'schen Apparat wurde ein bedeutender Gehalt an Dienkohlenwasserstoffen in diesen Fraktionen festgestellt. Da bei der Zersetzung des 1,2- und 1,3-Dimethylcyklohexans eine Bildung von Piperylen und Isopren möglich ist, wurden die niedrigsiedenden Fraktionen einer Bromierung und Kondensation mit Maleinanhydrid nach Diels und Alder unterzogen. Da wir jedoch nur über sehr unbedeutende Quantitäten dieser Fraktionen verfügten, gelang es uns nicht, bestimmte Resultate zu erzielen.

Experimenteller Teil

Gewinnung der Ausgangsprodukte. 1,2- und 1,3-Dimethylcyklohexan wurden durch Hydrierung des *o*- und *m*-Xylols (Kahlbaum) in einem Autoklaven in Gegenwart von Nickel gewonnen und hatten folgende Konstanten: 1,2-Dimethylcyklohexan—Siedepunkt 123.5—125.5° (bei 750 mm Druck, $d_4^{15} = 0.7902$, $n_D^{21.5} = 1.4310$; 1,3-Dimethylcyklohexan—Siedepunkt 118—120° (bei 750 mm Druck), $d_4^{15} = 0.7751$; $n_D^{21.5} = 1.4242$. 1,4-Dimethylcyklohexan wurde aus *p*-Methylcyklohexanon gewonnen. Mit Hilfe der Grignard'schen Reaktion wurde aus *p*-Methylcyklohexanon und Jodmethyl 1,4-Dimethylcyklohexanol-1 erhalten, das sich bei Destillation in Gegenwart von Jod zu 1,4-Dimethylcyklohexan-1 dehydratieren liess. Durch Hydrierung des letzteren über platinierter Kohle erhielt man 1,4-Dimethylcyklohexan, welches folgende Konstanten besass: Siedepunkt 118.5—119.5° (bei 750 mm Druck); $d_4^{16.5} = 0.7718$; $n_D^{21.5} = 1.4236$.

Apparatur und Arbeitsmethode. Die Zersetzung wurde in einer 80 cm langen Röhre aus undurchsichtigem Quarz ausgeführt, deren innerer Durchmesser 18 mm betrug und die auf einer Strecke von 40 cm mit Nichrom-Draht umwickelt war. Die Wickelung war mit einer Asbestschicht bedeckt. Die Temperatur wurde im Mittelpunkt eines erwärmten Raumes mittels eines Platin-Rhodiumthermoelementes gemessen, welches in einem Schutzrohr aus Quarz untergebracht war. Die Zersetzungsprodukte gelangten in mit Eis—Kochsalz gekühlte Behälter, wo sich die Hauptmasse der flüssigen Produkte ansammelte, und wurden darauf durch einen weiteren Behälter geleitet, der in ein Dewargefäss mit einer Mischung aus fester Kohlensäure und Äther getaucht war. Das Gas wurde

Resultate der Zersetzungsversuche mit Dimethylcyklohexane

Versuchsnummer	Substanz	Temperatur in °C	Versuchsdauer in Min	Durchgeleitete Menge der Substanz in g	Zuflussgeschwindigkeit der Substanz in ml/Min	Zuflussgeschwindigkeit des Wassers in ml/Min	Menge der erhaltenen flüssigen Produkte in g	Menge der zersetzten Substanz in g	Volumen des Gases unter normalen Bedingungen in l	Zusammensetzung des Gases in Volumprozenten							Ausbeute des Butadiens, bezogen auf die durchgeleitete Menge der Substanz in Mol %	Ausbeute des Butadiens, bezogen auf die zersetzte Menge der Substanz in Mol %
										CO_2	H_2	Gesättigte Kohlenwasserstoffe	Ungesättigte Kohlenwasserstoffe	Acetylenkohlenwasserstoffe	Propylen und Butylene	Äthylen	Butadien	
1	1, 2-Dimethylcyklohexane	675	99	36.1	0.5	1.8	27.8	12.1	4.81	1.0	15.8	35.8	47.4	0.6	14.2	23.2	9.4	6.3	18.7
2		700	74	29.7	0.5	1.9	15.8	20.2	8.24	0.5	18.4	31.1	50.0	1.5	14.4	26.8	7.3	10.1	14.9
3		725	70	27.3	0.5	1.9	16.2	17.6	8.13	0.6	16.2	32.7	50.5	1.4	12.8	27.3	9.0	13.4	20.7
4	1, 3-Dimethylcyklohexane	700	86	36.1	0.5	1.7	20.7	23.4	9.41	0.4	16.8	32.0	50.8	1.3	22.3	19.1	8.1	10.6	16.3
5		700	74	28.9	0.5	1.0	16.3	20.3	9.62	0.3	16.6	33.0	50.1	1.4	21.1	19.6	7.0	11.6	16.5
6		700	89	32.8	0.5	1.9	19.8	18.9	9.23	0.6	15.6	33.5	50.3	1.8	22.1	18.9	7.5	10.6	18.4
7		750	66	25.2	0.5	1.9	8.2	21.8	11.61	0.4	18.0	31.9	49.7	1.7	19.8	22.2	6.0	13.8	16.0
8	1, 4-Dimethylcyklohexane	700	71	24.9	0.5	1.8	17.2	11.9	4.94	0.7	17.0	29.7	52.6	0.9	24.1	16.8	10.8	10.7	22.5
9		750	57	21.7	0.5	2.0	7.2	18.7	9.73	0.6	16.8	32.2	50.4	3.1	18.8	19.3	9.2	20.6	23.9
10	Gemisch isomerer Dimethylcyklohexane	750	60	23.5	0.5	1.9	9.3	18.9	10.38	0.2	16.2	32.3	51.3	3.1	15.8	24.2	8.2	18.0	22.5

in einem Gasometer über Chlornatriumlösung gesammelt. Die Einrichtung des Apparates und Versuchsanordnung, sowie auch die Untersuchungsmethoden des Gases und des flüssigen Kondensats sind in der vorhergehenden Arbeit ausführlich beschrieben (1). Die Resultate der Versuche sind in der Tabelle zusammengefasst.

Nicht zulätzt möchten wir, dem Akademie-Mitglied, Prof. Dr. N. D. Zelinskij für das stete Interesse, das er dieser Arbeit entgegenbrachte, unseren ergebensten Dank aussprechen.

Zusammenfassung

1. Es wurde die thermische Zersetzung der 1,2-, 1,3- und 1,4-Dimethylcyklohexane in überschüssigem Wasserdampf untersucht.

2. Es wurde festgestellt, dass sich bei der Zersetzung von 1,2-, 1,3- bzw. 1,4-Dimethylcyklohexan Butadien bildet, jedoch in weit geringerer Menge als aus Cyklohexan und seinen monosubstituierten Homologen.

Aus 1,2-Dimethylcyklohexan wurde bei 725° Butadien mit einer Ausbeute von 20.7 Mol % (bezogen auf den zersetzten Kohlenwasserstoff), aus 1,3-Dimethylcyklohexan bei 700° 18.4 Mol % Butadien, und aus 1,4-Dimethylcyklohexan bei 750° 23.9 Mol % Butadien erhalten, während aus Cyklohexan und Äthylcyklohexan circa 40 Mol % Butadien gewonnen wurden (1).

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ZITIERTE LITERATUR

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BIOCHEMISTRY

THE PROBLEM OF BIOCHEMICAL CHARACTERISTICS OF VARIETY IN VEGETABLES

By B. A. RUBIN and L. I. NAUMOVA

(Communicated by A. A. Richter, Member of the Academy, 4. XI. 1935)

The result of the first investigations of our laboratory was the assertion that the activity of certain ferments may be considered as a relatively stable varietal character (1), to be specific, a connection was discovered between the energy of the development of the plant and the catalase indices. The activity of the latter in early varieties of radish proved to be considerably higher than in the late ones.

The task of further investigations consisted in finding the relation between the indices of the ferment system of the variety and its productivity, which here is meant as the energy of the formation and accumulation of certain plastic substances per unit of raw mass of the plant.

In investigating this question with regard to radish (2 varieties were taken, Erfurt, an early one, and Graivoron, a late one) we obtained the results set forth below.

The relation of the activity of ferments in these varieties is shown in Table 1.

Table 1

Activity of Ferment in Early Variety of Radish in Percentages of the Late Variety

Date of observation	Catalase		Saccharase		Amylase	
	Leaves	Root	Leaves	Root	Leaves	Root
June 20	110	226	171	492	112	103
July 10	101	274	127	147	141	101
August 3	117	178	111	104	181	107
August 31	102	136	171	237	114	116
September 22	148	141	156	—	146	103

We note the greater activity of all the three ferments in the leaves and roots of the early variety, which is in good agreement with the results obtained during our work in 1933.

It is interesting that the difference in the activity of the catalase is more

pronounced in the early stage of development (an excess of 125 and even 170%).

In the course of the development of the plants the differences between the varieties become to a considerable extent smoothed away, and during the last stage of observations the predominance of the early variety amounted to 40%.

The activity of catalase in the leaves of the early variety was found to be increased throughout all the stages of development, but the excess was considerably slighter all the time. Throughout all the stages of development, saccharase and amylase were considerably more active in the early variety. Thus, at the first stage of development, the activity of the saccharase and the amylase of the early variety exceeded that of the late one by as much as 38 and 23% respectively.

By the end of the vegetative period the difference as regards saccharase and amylase was 56 and 46% respectively.

In the roots of the early variety the saccharase activity at the beginning of the development reaches enormous values, viz. 322%, and only towards the end of the vegetative period sinks to 137%. As to amylase, its activity in the roots of the early variety is only slightly higher than that in the late variety. Besides differences in the activity of the ferments, essential differences between the varieties were established in the rate of growth of the raw mass and the accumulation of sugar.

The figures are shown in Table 2.

Table 2

Date of observation	Average weight in g				Total sugar in % of raw weight	
	Erfurt radish		Graivoron radish		Erfurt radish	
	Leaves	Roots	Leaves	Roots	Leaves	Roots
June 20	15.8	—	15.2	—	0.60	—
July 10	106.0	37.2	121.5	22.6	0.88	4.05
August 3	170.5	200.2	395.2	259.9	1.33	4.65
August 31	229.3	500.0	400.0	708.7	1.12	4.61
September 22	238.0	716.0	590.0	880.0	1.42	6.31

Date of observation	Total sugar in % of raw weight		Quantity of sugar per one plant in g			
	Graivoron radish		Erfurt radish		Graivoron radish	
	Leaves	Roots	Leaves	Roots	Leaves	Roots
June 20	0.37	—	0.0948	—	0.0562	—
July 10	0.80	3.19	0.9328	1.3041	0.9720	0.7209
August 3	1.52	4.61	2.2679	9.3093	0.0070	11.9814
August 31	0.97	4.65	2.5682	23.050	3.880	32.9545
September 22	1.61	6.44	3.3796	45.1796	9.916	56.6270

The table shows that, throughout all the stages of development (except the initial one), the weight of the plant of the late variety is always higher than that of the early variety

In the course of the development of the plants, the absolute sugar content, as well as the percentage of sugar, increases both in the leaves and in the roots, besides, towards the end of the vegetative period the increase of sugar stored up in the leaves of the early variety lags considerably behind the energy that characterizes this process in the late variety.

Thus, for instance, the sugar stored up in the leaves constituted the following percentage of the sugar stored up in the whole plant:

	Erfurt radish	Graivoron radish
August 31	10.0	10.5
September 27	6.9	13.7

We see that during September the percentage of sugar stored up in the leaves in relation to that stored up in the whole plant decreased by $^1/_3$ in the early variety and increased by the same amount in the late one.

Here we have indications that the Erfurt radish is nearer to the end of the firts year of development and to the accumulation of substances in the reserve organs and that, consequently, it is the earlier variety.

No essential differences as to the percentage of sugar are to be observed in the varieties under investigation. As to the sugar stored and the rate of accumulation of same, the differences are, on the contrary, very considerable.

In the first half of the vegetative period, the intensity of the increase in the stored up sugar is higher in the early variety, while in the second half it is higher in the late one.

At the same time, the late variety is considerably ahead of the early one in the energy of the development of the raw mass.

If we now compare the weight of the raw mass with the amount of carbohydrates stored up in the whole plant, we get the following figures (Table 3)*.

Table 3

	June 20	July 10	August 3	August 31	September 27
Weight of raw mass	105	100	57	66	70
Stored up sugar	171	132	65	70	74

The curve showing the relation of the varieties with regard to the amount of stored up sugar is at all the dates of observations higher than that showing the relation of the varieties with regard to the weight of the raw mass.

In other words, in the early variety of radish, the deposition of assimilates in the form of soluble, or in any case easily hydrolized, sugars predominates, as compared with the late variety, while in the late variety there predominates their conversion into cellular tissue, leading to the formation of new tissues and to an intenser increase of the mass of the plant.

We obtained analogous relations for two varieties of carrots, «Nantes» and «Valeria», of which the former is an early variety and the latter a late one.

The average activity of the ferments in the early variety in percentage of that in the late one is as follows:

	Catalase	Saccharase	Amylase
Leaves	112	137	137
Roots	119	111	162

* In calculating the figures, the data for the early variety were calculated in percentages of those for the late one, these latter being taken as 100.

It should be noted that not a single case was observed wherein the activity of a ferment in the early variety was lower than that in the late one. The relation of the stored up sugar also corresponded to the regularity noted for the radish.

Thus, it appears that the differences between the varieties as regards the energy of carbohydrates accumulation are connected, not so much with the different intensity of thier formation, as with the further fate of the assimilates formed.

Consequently, the utilization and further transformations of the initial assimilate take different course in the different varieties. There is no doubt that the above-mentioned differences are also connected with the differences in the fermentative system.

The investigation of these differences, which involves great difficulties is an extraordinarily interesting task. It should be noted that this problem is of particularly great importance with regard to the sugar beet, as the causes of the negative correlation between the productivity and the sugar content of this plant, which has been confirmed many times, remain quite unknown.

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BIOCHEMISTRY

THE CONSUMPTION OF SUCCINIC ACID BY FUNGUS FILMS AS AFFECTED BY THE ACETATE

V. S. BUTKEVIČ (BUTKEWITSCH), Corresponding member of the Academy, and L. K. OSNICKAJA

It has been shown by the experiments of various authors (1) that, if certain fungi are placed on acetate solutions, an accumulation of succinic acid in such solutions may be observed. Such observations as were made have so far been construed in the sense that succinic acid is formed from acetic acid through the dehydration of the latter. It is this construction put on the observations that is responsible for the development of a general scheme of the biochemical formation of succinic acid out of sugar via acetic acid, involving a prior splitting up of its molecule similar to that which takes place in alcoholic fermentation (2).

A similar scheme was later extended to the biochemical origin of sugar out of citric acid, the starting point also in this case being observations of the accumulation of that acid by certain fungi when their films were placed on acetate solutions (3). A more detailed study of this accumulation of citric acid carried out by one of the writers, in cooperation with Menzschinskaja and Trofimova, has shown that citric acid does not, in these cases, owe its origin to acetic acid, but is produced by substances contained in the mycelium, so that the accumulation of the acid so formed on the solution of the acetate is merely due to the protective action exercised by the latter with regard to the acid in either doing away with or reducing its consumption by the fungus (4).

After the protective action of acetates with regard to citric acid had been proved, the question naturally arose as to a similar action exercised by them on succinic acid. It was to solve this problem that the experiments described below were made by the writers.

The experiments were arranged in the same manner as those carried out with a view to ascertaining the influence of the acetate on the consumption of citric acid by *Aspergillus niger* films (4).

Two series of experiments were made. Three-day films of *Aspergillus niger* (strain 1a), grown on 100 cm³ of the usual medium (with 5% of sugar and 0.3% of NH_4NO_3), at a temperature of 30°, were used in either series. The solutions tested were kept covered with the films, at a temperature of 30°, for 5 days in the first series of experiments, and for 2, 4, or 5 days in the second. In all cases, the amount of sodium acetate introduced into the test

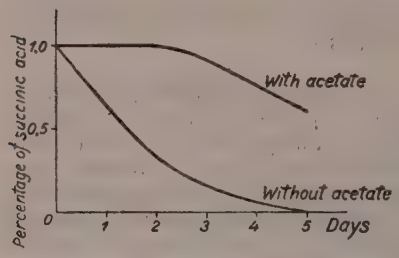
solutions was one corresponding to 2.5% of acetic acid, while the amount of sodium succinate corresponded to 1% of succinic acid.

The succinic acid content in the test solutions was determined quantitatively, by extracting it with ether out of the acidulated solution, after the oxalic acid had been removed from the latter, and by titrating the aqueous solution left after the evaporation of the ether extract with $N/10$ alkali. The quantitative determination of acetic acid, citric acid and sugar was made in the same manner as in the experiments with citric acid ([4]).

The results of the experiments (averaged from two parallel tests) are given in the table, which also shows the initial composition of the test solutions.

The variation of the consumption of succinic acid in the 2nd series of experiments, both with the acetate and without it, are graphically shown in the appended diagram.

In the experiments made by the writers, the acetate exhibited the same well defined protective action in relation to succinic acid as it did in the experiments mentioned above with regard to citric acid. So long as the acetate content of the solution is high, practically no consumption of succinic acid takes place, the latter becoming noticeable only after no more acetate has been left in the solution. Whereas in those experiments of the 2nd series that were made without the acetate, nearly the whole of the succinic acid introduced into the solution was consumed before the close of the 4th day, only 25% of its initial amount was found to have been consumed within the same period of time in the cases where the acetate was used, in spite of the fact that the latter had been almost wholly exhausted.



In the concentrations tested (1—2%), sugar failed to exercise any appreciable influence on the consumption of succinic acid (no more than it affected that of citric acid in previous experiments). A more complete consumption of succinic acid by a test solution containing 2% of sugar than by an unsugared one as observed in the first series of experiments, may have been due to a more vigorous growth of the film on sugar.

In either series of experiments, no citric acid could be detected, unless some sugar (1—2%) was added to the acetate. As may be inferred from our former experiments, this accumulation of citric acid on dilute sugar solutions was due to a reduction in the consupmtion of the acid, which had in this case been formed from the sugar introduced into the test solution.

As regards the accumulation of oxalic acid, it corresponds in most cases pretty accurately to the amount of the base liberated by the consumption of succinic and acetic acids from their sodium salts.

The ptotective action of the acetate on succinic acid, as ascertained by our experiments, involves a revision of the construction formerly put on the evidence regarding the accumulation of succinic acid on acetate solutions. It is possible that, in those cases where such an accumulation was observed, the succinic acid had been formed out of some substances contained in the mycelium, and not out of the acetate, the latter merely protecting the acid derived from some other source. Further research is needed to elucidate this question, and, in the meantime, the scheme of the formation of succinic acid out of sugar via acetic acid, involving the splitt-

Test solution (100 cm³)	Duration of experiment (days)	Acidity in cm³ $N/10$ NaOH per 10 cm³	Succinic acid	Acetic acid	Oxalic acid	Citric acid	Sugar	Weight * of film in grams
			Percentage in solution					
1st Series of Experiments								
Sodium succinate	5	6.9	0.24	—	0.87	0	—	0.43
Sodium succinate plus sodium acetate	5	Alkaline	0.98	0.54	1.00	Traces	-	1.04
Sodium succinate plus 2% sugar.	5	12.0	0.05	—	1.05	0	0.04	1.83
2% sugar plus sodium acetate	5	1.8	—	0.30	1.35	0.22	0.22	1.26
2nd Series of Experiments								
Sodium succinate	2	4.1	0.31	—	0.82	0	—	1.33
	4	4.3	0.05	—	0.89	0	—	0.89
	5	3.8	0	—	0.84	0	—	0.79
Sodium succinate plus sodium acetate	2	Alkaline	1.01	0.66	0.94	0	—	1.04
	4	2.0	0.74	0.05	1.72	0	—	0.79
	5	5.4	0.60	0	2.27	0	—	0.92
Sodium succinate plus 1% sugar.	2	7.4	0.37	—	0.76	0	0.26	1.66
1% sugar plus sodium acetate	2	Alkaline	—	1.23	0.44	0.16	0.24	1.09
	4	Alkaline	—	0.67	0.89	0.17	0.13	0.93
	5	5.2	—	0	1.50	0	Traces	0.99

* The average weight of the film originally used in the test was 1.88 g in the 1st series, and 1.52 g in the 2nd series of tests.

ing up of the sugar molecule similar to that taking place in alcoholic fermentation, must be recongnized as not sufficiently warranted by evidence.

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BIOGEOCHEMISTRY

VANADIUM IN THE PETROLEUMS AND BITUMENS OF THE USSR

By A. P. VINOGRADOV and G. G. BERGMAN

(Communicated by V. I. Vernadskij, Member of the Academy, 17. XII. 1935)

Many of the rarer chemical elements—Cu, Ag, Ni, Mn, Pb, Cr, U, Mo, V, Ti, As, Li, Ba, Sr as well as others—have been found repeatedly in the ash of petroleums, asphalts and other bitumens. Some of them appeared to recur constantly in petroleum with a certain degree of concentration, for example nickel. The presence of vanadium in comparatively large amounts in some of the petroleums and bitumens is, however, of greater significance. In this respect solid bitumens differ from coal, the only exception being cases of the infiltration of coal by solutions containing V; but under these circumstances the coal usually contains other heavy metals also, *e. g.*, U, Cr, Cu and so forth).

It may now be safely asserted that the presence of V in petroleum and bitumen is not an accidental or local phenomenon but has a more general, geochemical character.

Briefly speaking, it may be said that a high content of V was first observed in South America in the Argentine bitumen known as rafaelite and subsequently in other bitumens and petroleums of that country also (*v.* Longobardi and Camus, 1911; Longobardi, 1934; Fester and Bertuzzi, 1925, 1934, *et al.*); later it was observed in the petroleums and solid bitumens of Peru, for example in Yauli region (*v.* Hewett, 1910; Baragwanath, 1921); and, finally, in the petroleums and bitumens of Venezuela, such as Trinidad asphalt (*v.* Salkover, 1910, *et al.*), in the petroleums of Lake Maracaiba, and elsewhere (*v.* Broz, 1930, *et al.*). The presence of V is very probable in other South American petroleums, which are, in this respect, of unusual interest. Vanadium has been found in various bitumens of the United States, known under different names (*v.* Hillebrand, 1910; Wells, 1925; Nason *et al.*), and also in some of the petroleums (*v.* Shirey, 1931; Bird and Callot, 1930, *et al.*). Considerable amounts of V are known to exist in the petroleums of Eurasia,—of Hungary (*v.* Sürü, 1935), Persia (*v.* Dunstan, 1924; Thomas, 1924, *et al.*), and of Rumania. And, finally, it is known that V occurs in the petroleums of Egypt and other places (*v.* Guthrie, 1923).

In the Soviet Union V has been found in Velikhov bitumen (*v.* N. A. Orlov, 1932). Nothing has been known about its presence in USSR petroleums.

The existence of marine organisms which concentrate vanadium allowed one of us, at one time, to advance the theory of a connection between

Vanadium Content in Petroleum and Others Bitumens

Name	Origin	Period	Ash content %	U content in percentages of ash	V content in percentages of bitumen
Nova Zembla, asphaltite	Nova Zembla, Cape Sokol, right lens	Devonian	3.44	3.1	$1.06 \cdot 10^{-1}$
» » »	» » » » left »	«	4.80	2.45	$1.17 \cdot 10^{-1}$
Ukhta petroleum	Ukhta River, Chibyu	Upper Devonian	0.0387	9.60	$3.72 \cdot 10^{-3}$
Izhma asphaltite *	Ukhta region, bank of the Izhma River	» »	0.140	28.37	$4.34 \cdot 10^{-2}$
» »	» » » » » » »	» »	0.235	17.85	$4.02 \cdot 10^{-2}$
Chusovaya petroleum	Chusovskie Gorodki	Permo-carboniferous	0.0041	4.00	$1.64 \cdot 10^{-4}$
Sterlitamak »	Ishimbayevo (bore-hole 733, depth 507 m)	Permian	0.0120	28.63	$3.44 \cdot 10^{-3}$
» »	Ishimbayevo (bore-hole 733, depth 507 m)	»	0.0447	12.7	$5.50 \cdot 10^{-3}$
Syzran, asphalt	Middle Volga, Syzran, Samarskaya Luka, adit	»	0.42	12.5	$5.25 \cdot 10^{-2}$
Kairovka asphalt	Kairovka Village, near Orenburg	Upper Devonian	0.07	2.89	$1.96 \cdot 10^{-3}$
Emba petroleum	Koschagyl, bore-hole 8	Jurassic	0.0131	0.247	$3.24 \cdot 10^{-5}$
» »	South Iskine, bore-hole 7	»	0.00243	0.327	$1.00 \cdot 10^{-5}$
» »	« » » «	»	0.0352	0.102	$3.60 \cdot 10^{-5}$
Surakhany petroleum	Baku region Surakhany	Tertiary, pliocene	—	Absent	—
Grozny »	Grozny	» miocene	0.010	0.781	$8.22 \cdot 10^{-5}$
Coke	Baku petroleum		1.36	0.1	—
Shimakha asphalt	Caucasus, Shimakha		3.98	Traces	—
Ozokerite	Krasnovodsk region		0.29	Absent	—
»	Borislav, Galicia		1.83	Traces	—
	Other Bitumens				
Anthraxolite	4 km from Shira Sta. SW Aginsk-Min. R.R. (Melaphite bed)	Permo-carboniferous	7.04	0.030	$2.16 \cdot 10^{-3}$
Asphaltite	From Dzungarian border **		0.22	0.15	$3.30 \cdot 10^{-4}$
Gilsonite	American		0.19	0.12	$2.42 \cdot 10^{-4}$
Gagat (Pitch coal)	Unknown		2.30	0.14	$3.22 \cdot 10^{-3}$
Trinidad asphaltite	South Amer., from Lake Trinidad petroleum	Upper Cretaceous	36.50	0.047	$1.71 \cdot 10^{-2}$

* In the other sample of ground asphaltite containing 0.9 ash was found $2.42 \cdot 10^{-2}$ in asphaltite or 1.32 per cent V in ash.

** See the description in the paper of B. A. Obruchev, «The Dzungarian Border».

the origin of V in petroleum and such concentrator-organisms (1). The author made an attempt at that time to find facts which might indicate exactly in what way the concentration of V in petroleum and solid bitumen took place, and was successful in proving that there are a considerable number of existing marine organisms, *viz.*, the ascidians, which concentrate V. However, it was found much more difficult to determine the moment at which the V actually entered into the petroleum. An elucidation of the laws governing the distribution of V, not only in organisms, but also in petroleum, was requisite.

The material on the presence of V in petroleum so far available cannot be considered adequate to solve the problem. Very often, for instance, in the works we have cited, no information is given as to the origin of the petroleums, bitumens, *etc.*, or of their treatment. This led us to undertake a systematic study of the natural petroleums and bitumens of the Soviet Union with respect to the peresence of V, the results of which, in our opinion, may have not only a scientific interest but perhaps also some practical value (*e. g.* the possibility of using the V from petroleum ash in industry). For this purpose, we decided to make a more detailed study of various petroleums which are rich in V.

The determination of V in such petroleums was done in the following way. The petroleum (500 cm^3) was evaporated, a continual stream of air being forced through it. The solid residue was then carefully burned in porcelain crucibles, and the ash thus obtained dissolved by fusing it with $Na_2CO_3+NaNO_3$. The V was precipitated from this solution in the form of mercury vanadate and then determined by means of phospho-tungstic acid by a method which we have described elsewhere (2). We give below the figures for the average content of V in the crude petroleums and natural solid bitumens of the Soviet Union, taken chiefly from beds located approximately on the meridian running through Nova Zembla, Ukhta, the western slope of the Urals, the Emba region and, finally, the Caucasus.

It follows from these data that the petroleums of more ancient origin contain the most V. The petroleums, and the solid bitumens associated with them (*e. g.*, in Nova Zembla and Ukhta), belonging to the Devonian deposits, contained the largest, amounts of V. This is also true for the petroleums of the Chusovskie Gorodki and of Sterlitamak belonging to the Permo-carboniferous deposits. On the other hand, the petroleums of the Emba region, of the Jurassic age, and those of the Caucasus (Tertiary), as a rule contained no V or very little. The concentration of V in the bitumens of the Devonian period has a geochemical interest, since it indicates a considerable migration of V at that time (the Devonian period), its fixation in sea mud and so forth. The presence of V in the Devonian petroleums of the Soviet Union does not justify us in assuming that this would hold true of all Devonian petroleums. On the other hand, it does not exclude instances of the accumulation of V in the petroleums and bitumens associated with other ages. For example, the South American petroleums which are high in V are associated with the Cretaceous deposits. Therefore the interpretation of these facts is of special importance in the geochemistry of certain regions and especially, in our case, for the Devonian and Permo-carboniferous rocks of the Northern part of the USSR and the Ural region.

From much of the data, *e. g.*, those for the Izhma asphaltite and some others, it appears that the content of V is in proportion to the organic matter, but not to the ash of the rock. Our petroleums which are high in V are the heavy petroleums with an asphalt base and containing considerable

amounts of sulphur; fot instance, the petroleums of Ukhta, Sterlitamak and the Chusovskie Gorodki, while the paraffin petroleums, for instance those of the Caucasus and Emba, contain little V or none. Moreover, we were unable to find any V in a number of ozokerites. These results correspond with the investigations made of European and American petroleums, in which the higher content of V in the asphalt petroleums containing large amounts of sulphur, was also noted. This fact led some investigators to assume that V passes over into the petroleums from aqueous solutions coming into contact with them. The presence of H_2S causes V to pass over into V_2S_5 and in this form it is retained in the petroleums. However, this theory is now regarded as unfounded. The possibility of the existence in the past of natural overground waters containing V cannot be denied, but we must bear in mind that the water associated with petroleum is usually salt water, derived from sea water (and in many cases probably from the fossil mud waters of marine bottoms)([3]). Even if such waters were in some way, enriched by V (which is unlikely), their contact with the petroleum is quite insufficient to account for the transfer of V (in the form of V_2S_5?!) to the petroleum.

At one time there was another theory of the origin of V in petroleum, *viz.* that it came from the mud of continental sea basins in which V concentrated partly through such marine organisms as the ascidians ([1]).

The basis of this point of view was the idea that V is not introduced into the completely formed petroleum with the water, but that it has been present in the petroleum from the moment of its formation, and in this way, owing to its catalytic properties, exercises an influence on the ultimate character of the petroleum formed. Specifically, therefore, the V in the petroleum may occur not only in the form of non-organic compounds, but also of organic, and, as a matter of fact, Treibs ([4]) has already demonstrated that in a number of petroleums V appears in the form of a porphyrine complex. Under these circumstances the porphyrine has obviously been formed from chlorophyll. Without at present going into the details of the process of the migration of V into the petroleum, we wish to note the fact, since it supports the point of view referred to above.

Besides the petroleums and bitumens of Ukhta, the western Urals, Emba and the Caucasus, we have also quoted some data regarding the determination of V in other bitumens of the Union. Further research along these lines will help to solve the problem we have posed.

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GENETICS

KARYOLOGICAL INVESTIGATION OF THE EMBRYONAL OOGENESIS IN MAN

By A. H. ANDRES and J. VOGEL

(Communicated by A. A. Sapehin, Member of the Ukrainian Academy, 4. X. 1935)

The present investigation constituted a karyological study of the nucleus of the germ-cell in the human embryonic ovary. In particular, it aimed to check upon the findings of Evans and Swezy (5). The latter reported that germ-cells of embrional origin degenerate and that definitive ova of adults are formed *de novo*, independent of the embryonal ones and only during the period of sex-puberty. Eleven fetal ovaries aged 2½ to 8 months were studied; six of them (up to 6 months of age) were obtained by means of laporotomia, immediately dissected, and the gonads fixated. Embryos aged 6 months and older, some of them stillborn, were dissected immediately, and their gonads fixated; some prematurely born who had lived for some time were dissected instantly after death. Fixation was carried out with strong solution of Flemming's fluid, according to Painter's method and Navashin's fluid. Paraffin embedding. Sections 10 and 15 μ thick. Stained according to Michaelis, Newton, Heidenhain, Feulgen.

Conclusions

1) Processes of reproduction and development of germ cells reaching the stage of zygotene were observed during embryonal oogenesis in man. Embryonal cells that remained after the formation of the primordial follicle degenerated, showing peculiar pictures of picnosis of separate chromosomes. The other cells that enter the follicle undergo no degeneration during emryogenesis. It cannot be said, therefore, that complete degeneration of embryonal generations of germ-cells occurs during embryogenesis.

2) There was observed in some cases, already during the embryonal period, activation of several primordial follicles and their development into almost a «mature» Graafian follicle. The germ-cell nucleus enters this period of growth and maturation in the stage of zygotene and remains in this state until the time of maturity of the follicle. During the process the, germ-cell itself and its nucleus become enlarged (as does the whole follicle), the chromosomes of the germ-cell become swollen, they are not as well stained, and the Feulgen's reaction becomes reduced. These changes should be considered as the great period of growth of the human germ-cell.

3) No maturation divisions take place in the embryonic ovary.

4) Polyovular follicles are sometimes present in man. They are formed either from the penetration of accessory oocytes into the primordial follicle or from the several accessory somatic mitoses. In this case the polyovular follicle is formed by an oogonium separated from other oogonia, before it is surrounded by follicular cells.

5) The chromosome set of an oogonium consists of 48 elements. Among these are 10 individualizable autosomal pairs and 2 *X*-elements.

6) Together with such orthoploid cells, a certain insignificant number of oogonia show polycentric mitoses and various destructions in their course (lagging chromosomes and their elimination, heterokinesis) leading to the formation of heteroploid sets.

7) It is probable that tetraploid germ-cells may be formed in the course of oogenesis just as it takes place during spermatogenesis in man.

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GEOLOGY

ON THE AGE AND THE CONDITIONS OF FORMATION OF BROWN COALS IN THE UPPER ZEYA PLAIN

By G. E. BYKOV

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In the course of the geological investigations performed by an Expedition headed by A. N. Churakov, the writer succeeded in discovering brown coal deposits at various points of the western part of the so-called Upper Zeya plain, which he had been commissioned to study.

The most promising outcrops are those of Chorny Kluch (Blac Brook), along Dzhamkuchi river (a tributary of the r. Tunylghin). Here, one of the outcrops reveals a series of alternating layers of brown coal, mud clays and grey sands, overlaid with some 9 meters of obliquely stratified quartzsands containing in their lower portion narrow (0.5—1 cm thick) bands of coal. The fossil strata have a thickness of 0.12 m (upper stratum), 0.25 and 0.32 m (lower stratum). Lower down along the brook, a number of outcrops make it possible to obserbe the gradual disappearance and tapering off of the upper two strata. As to the lower, its practically undiminished thickness (0.30 m) is revealed by a cliff 75 m distant from the outcrop described above.

Other outcrops of brown coal have been located by the writer in the lower reaches of Tunylghin river, near a group of lakes that are nameless. In one of the numerous hills occurring here and due to ice pressure, the inner structure of which has been revealed by detrition, the mass of ice constituting the body of the hill is seen to contain fragments of brown coal of various sizes mixed up with grey clay. In the very roof of the «bolgunyakha» (a local name for large hills of that description raised by ice pressure), there have also been found several bands of a black mass made up of fragments of coal mixed with sand and small pebbles. Owing to the deformation produced by the raising of the hill, the layers of pebbles, as well as the embedding sands, are inclined to the horizon at varying angles up to 35°. These facts indicate the occurrence at this point, at a shallow depth below the surface, of one or more thick layers of brown coal. Smaller outcrops of the latter are to be seen along Timnya river and its tributary, the Dessu, as well as along Golovitkan river and elsewhere. Abundant evidence of their occurrence in boreholes and in the sediments of some brooks studied by the writer and other investigators (1) would warrant the assumption that brown coals are widely developed here.

The brown coals of this district show the following physical characteristics. They are either brown or fulvous in colour, the shades varying from stratum to stratum. The ash left after burning is also of a pale-brown colour.

In a sample drawn from the Dzhamkuchi deposit, its content was 5.6%. Percentages of other substances: hygroscopic humidity, 17.6%; sulphur, traces; volatile substances, 49.1; coke 27.7. The coal burns with a long sparkling flame. Fragments of compressed wood, scales of fruit cones of conifers, etc., are visible to the naked eye among the amorphous mass of the coal. Of special interest are inclusions of charred bits of wood and even of tiny particles of compressed charcoal. The presence of the latter is probably due to the fact that at the time of the deposition of the coal measures, frequent conflagrations took place, the charred trees being buried in the layer along with other organic matter. The specific weight of the coal is 1.44. The surface of the coal is opaque with occasional dark lustrous patches. The fracture is rugged. The coal is very hard to break. Its chief characteristic is the frequency of coarse fissures and its perpendicular lamination, the parallel planes of its fissures being, to judge from the outcrops, strictly oriented.

The brown coals occur in the lower part of a series formed with the detritus of the Soktakhan mountains* and which may therefore aptly be termed the Soktakhan series. It overlays conformably the granite-gneiss bed of the Upper Zeya plain, and rests partly on Jurassic deposits, with which it exhibits a sharp angular unconformity. The series may conditionally be subdivided into two portions: the lower one, made up of mud clays, brown coals, sands and occasional pebbles, and the upper, arenaceous one. The latter is made up of sands of varying grades of coarseness, with an admixture of well rounded pebbles, more abundant in the upper layers and more scarce in the lower ones. These formations also exhibit a coarse and sometimes oblique stratification and offer an unusually heterogeneous alternation of lenses and bands of sands, sandy clays and pebbles. The series shows a general northward dip, equal to an incline of 0.0015. It is of varying thickness, 100—150 m on an average. Contiguous to that series are quarternary deposits found in river valleys. They also form a thin cover on watersheds, concealing the underlying strata of the series.

The age of the Soktakhan series is determined by the composition of the pollen contained in the brown coals. The determinations of the latter made by E. Smirnova (some of which appear on the appended diagram) show them to contain vegetable remains of a type differing widely from that to be found in the district at the present time. In addition to the forms that have survived here to the present day, such as *Betula*, *Pinus*, *Alnus* and *Picea*, the fossil coals contain a number of forms utterly foreign to the present-day vegetation of the district. Such are: *Tilia*, *Juglans*, *Fagus*, *Ulmus* and others. *Larix*, which is a widely distributed form at the present time, has not been identified in the coals. On the other hand, the coals contain such forms as *Pterocarya*, *Ilex*, *Tsuga*, *Fagus*, whose range of occurrence is very restricted at the present time (*Tsuga*, for instance, is only encountered in some places of North America and South-Eastern Asia), but which were widely distributed in the Tertiary Period, as shown by A. N. Krishtofovich (2). The type of vegetation described has been repeatedly mentioned in literature as characteristic of tertiary deposits. The vegetable remains identified in the coals are most closely related (if the comparison be limited to fossil types of the Far East) either to the Upper Dui vegetation in the Sakhalin Island or to the fossil flora of the coal measures round Korff Bay, which have been referred to the Oligocene or the Lower Miocene. The only point of difference is

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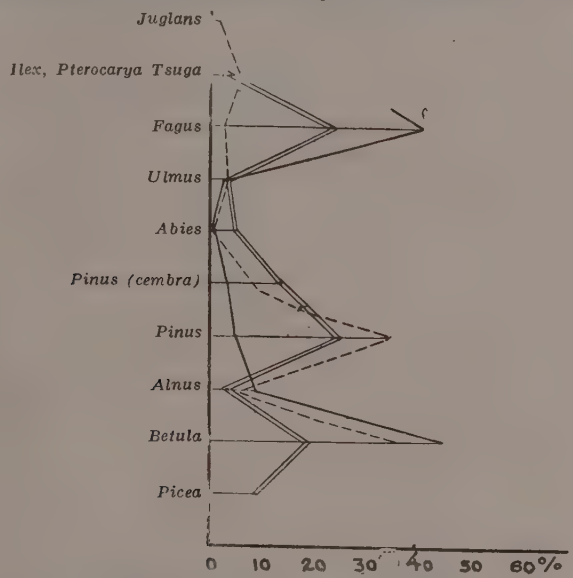
* This is the name given locally to the part of the Tukuringra range that lies between the Zeya river and Ogoron Lake.

that the Soktakhan flora contains a higher percentage of conifers. All this would suggest that both the brown coal and the whole of Soktakhan series (containing in its upper layers fossil wood which, according to Ermolenko's determination, also belongs to thermophilic plants) might be referred to the Neogene, without, however, assigning them for the time being to any particular horizon.

It is to be noted that the Soktakhan series, and its brown coals in particular, exhibit certain traces of changes due to dislocation. Thus, the conditions of occurrence of the strata revealed by the Dzhamkuchi outcrops show



Composition of the pollen contained
in the brown coals of the Soktakhan series.
——— Brown coals from Chorny Kluch (lower stratum).
═══ Brown coals from Chorny Kluch (upper stratum).
– – – – Brown coals from lower reaches of Tunylghin river.

them to have a slight incline, equal to 6—7°, with an azimuth of dip of 170° to the south east. The dip of the strata ought not to be confused with either the oblique lamination or with the incline of the basin, since both the lamination and the tapering off of individual lenses are already to be observed within the dipping strata. Other traces of dislocation are the fissures observed in the brown coals, which, in some rare cases, show a displacement of the seam of 1—2 cm along the fissure. The planes of fissure observed in the outcrops on Dzhamkuchi river were strictly parallel to one another. Two planes are clearly distinguishable, their strike being 20° and 85° NE, and their dip approaching the vertical. These phenomena are observable in the brown coals in spite of a total absence of any traces of cementation in the sands in which they occur. The clays in which the coals are embedded also exhibit only a very slight degree of cohesion. Only on the Dessu river do the overlying sands show some vestiges of cementation, visible from the fact that blocks of rock may be detached from the solid, a light blow causing the latter to split up into sheets along the planes of stratification, instead of reducing it to powder. Still, the cohesion of the rock is but slight even at this point.

The general structure of the series is practically unaffected by the dislocation observed in the brown coal strata, since, as mentioned before, the former rests conformably on crystalline rocks, with a general gentle northward dip.

The conditions of sedimentation of the rocks containing the brown coals do not seem to have been different from those observable in river sedimentation at the present time. This is evidenced by the lithological composition of the series (nearly all sands), the absence of finely sedimented deposits, a coarse and irregular oblique stratification, a frequent petering out of bands and lenses. Palaeogeographic considerations suggest the existence in the past of a great river, flowing in a latitudinal direction and receiving from the south a number of small tributaries, which rose in a high hilly watershed. In this respect, the Soktakhan series is somewhat similar to the gigantic cone of deposits formed with the detritus of the Soktakhan hills. The coal bands themselves are due to the deposition which took place in the silted-up river beds, marshes and lakes lying in the flood plain of that river. This is shown by the fact that the brown coals, as well as the mud clays associated with them, occur in the shape of lenses, and also by the smallness of the areas taken up by individual layers. The lakes deposited at their bottom the organic remains of the surrounding luxuriant forests, and, as a result of floods, were ultimately choked up with sand and silt, their sediments being buried under the latter. Such conditions of sedimentation could hardly result in the formation of coal strata of any commercial value, still, the frequent recurrence of the conditions to which they owe their accumulation in the Upper Zeya plain would make probable the occurrence of small layers or lenses of coals over considerable distances. It is not impossible that among the areas they occupy some may be found to contain somewhat larger units, which would warrant future mining operations.

As shown by the list of arboreal vegetation the pollen of which has been preserved in the coals, the climate prevailing at the time of the deposition of the coal strata was a warm, but not a hot one.

A few words as to the conversion into coal of the organic remains accumulated in the basins. The most essential changes in the sediments have admittedly occurred in the course of those dislocations which gave rise to the dip of the strata and the fissures described above. Although the pressure of the overlying masses did play a certain part, it had never been a factor of any importance, since the rocks resting on the brown coals had never in the past exceeded a mere 30—50 m in thickness, at a rough estimate. A small remnant of the arenaceous series still surviving as Chorny Kluch Hill situated near the Chorny Kluch deposit, has its summit at only 30 m above the coal strata level. That maximum height is not exceeded by any watershed in the plain. This being the case and in the absence of any evidence to the contrary, it should be admitted that the dislocation of the brown coal strata is due to unilateral pressure from the sides, combined with the relatively small weight of the overlying sands. Although we possess no data that would unambiguously testify to the direction of the pressure, still it is to be presumed that it came from the side of the mountain ranges bounding up the Upper Zeya plain in the north and in the south. It must have been a meridional one. This assumption would be in good agreement with the above-quoted measurements of fissures and of the dip of the brown coal strata.

In conclusion and by way of preliminary inferences it may be stated that the Upper Zeya plain has some characteristics bearing a resemblance to D. V. Nalivkin's «continental geosynclines». Its situation along the zone

of recent risings in the Tukuringra—Soktakhan—Dzhagdi ranges, the great thickness of the alluvial deposits, the traces of dislocations in the brown coal seams, and, lastly, the incipient subsidences in the region of the plain resulting in the formation of buried terraces, all this would indicate that we have in this instance an incipient continental «geosyncline» zone, which began to form as late as the end of the Tertiary Epoch.

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PLANT PHYSIOLOGY

ON HASTENING THE RIPENING OF MELONS

By J. V. RAKITIN

(Communicated by A. A. Richter, Member of the Academy, 23. X. 1935)

Notwithstanding their numerous valuable qualities (high sugar content, tender delicious meat, agreable odour and finally their high productiveness), the melons of the Stalingrad district have, up to the present, only local importance.

The export of fruit products, even with cheap water transportation, bears a casual character.

The basic reason for the slight development of melon culture in the Stalingrad district is the inability of the ripe product to hold up under prolonged transportation.

A study of the conditions of melon cluture in the Stalingrad district and work concentrated on the theory and practical side of the ripening of fruits have proved that the difficulties of transport are fully conquerable provided the fruit be transported when quite grown but not yet ripe (in which state they can be kept during a very prolonged transportation) and brought to complete ripeness afterwards in the place where they are used.

The method of hastening ripening that was used was that of the American investigators Denny ([1, 2, 3]), Rosa ([4]), and Harvey ([5, 6, 7, 8]), a method based on the stimulating action of ethylene.

But inasmuch as there are no directions [with the exception of a short note by Harvey ([9])] in the literature on the ripening of melons, the technique of their ripening had to be worked out a new. As a result of repeated laboratory experiments it was established that the best concentration of the gas for the purpose of hastening ripening and for securing a high quality of product is one of 1:2000. After 4 to 6 charges of gas at this concentration, with a 24-hour interval after each charge, the fruit took on the usual qualities of the ripe melon. The control melons ripened no sooner that from 25 to 30 days.

At the same time, along with the working out of a technique of hastening ripening experiments were conducted to ascertain the biochemical essence of the stimulating action of ethylene. This had already been studied principally by the Americans Nord ([10]), Dykins ([11]), Englis and Zannis ([12]) Harvey and Regeimbol ([9, 13, 14]), Denny ([15, 16]), also by Russians Tserevitinov ([17]), Iva-

nov-Prokoshev and Gabunia [18], and others. But the investigations of all the authors enumerated were directed toward ascertaining the changes in activity of ferments, permeability of the plasm, and the like. All these give but an indirect conception of the effect of ethylene on the metabolism. It is to be doubted that the essence of the action of ethylene can be understood without a general picture of the transformation of substances. Hence, it has heen considered necessary in this, as also in my first work, finished in September 1934 [but passed to press only during the present year [19]], to carry out the investigation of the biochemical essence of the action of hastening ripening from the point of view of changes in metabolism. We have considered it possible to judge of the changed course of metabolism by the formation of acetaldehyde and ethyl alcohol.

The alcohol was determined according to Nicloucx [20], the acetaldehyde by the method described by Agabalyants aud Savenkov [21].

The first experiments were concerned with determining the dynamics of the formation of C_2H_5OH and CH_3COH by fruits at various stages of their development.

The results are shown in Table 1.

Table 1

Variety	Ripeness of fruit	C_2H_5OH in mg per 100 g of fruit	CH_3COH in mg per 100 g of fruit
Biriuche-Kutskaya	Green, not yet grown	11.5	0.10
	Green, fully grown	17.0	0.15
	Beginning to turn yellow . .	24.1	0.20
	Ripe	37.0	0.79
Cassaba spotted	Green, not yet grown	14.0	0.22
	Green, fullygrown	20.3	0.32
	Beginning to turn yellow . .	35.1	0.44
	Ripe	40.0	0.85

From these figures it is apparent that the metabolism deflects toward intensification of anaerobic processes according to the degree of ripeness.

The following experiments were concerned with a determination of the action of ethylene on the formation of C_2H_5OH and CH_3COH. Large, full-grown melons were chosen for the experiments. They were cut into two symmetrical halves. The cut areas were dried in the sun. After being so dried, the areas of the cuts were not infected by microorganisms during the course of 8 to 10 days. The melons, dried in this manner, were placed in chambers (tight wooden boxes of about 56 litres capacity). The method of using halves prevented mistakes that would have been possible in working with whole melons. These halves were subjected to the action of ethylene gas in a concentration of 1 : 2000 twice in 48 hours.

The results are shown in Table 2

Table 2

Variety	Conditions of Ripening	C_2H_5OH in mg per 100 g of fruit	CH_3COH in mg per 100 g of fruit
Biriuche-Kutskaya	Control	18.5	0.185
	Ethylene	25.8	0.291
Cassaba spotted	Control	20.1	0.292
	Ethylene	25.4	0.553
Cassaba wrinkled	Control	19.1	0.255
	Ethylene	27.3	0.396
Zimovka	Control	20.4	0.311
	Ethylene	24.6	0.521

The data obtained show that ethylene, while hastening the ripening of fruit, deflects the metabolism toward intensification of anaerobic processes.

The question arises: is not this deflection of metabolism one of the basic links in the chain of processes taking place in the living cell under the influence of ethylene? In other words, is not the partial anaerobiosis induced by ethylene, of which C_2H_5OH and CH_3COH are indicators, the essence (to a first approximation, at least) of the hastening of the ripening of fruits, and are not the products of anaerobic metabolism, viz. C_2H_5OH and CH_3COH, the cause of an intensified transformation of substances.

The results of the following experiments throw light on the questions raised.

The first experiment shows the possibility of hastening the ripening of fruits by the injection of alcohol.

The concentrations of alcohol used were the following: 48, 24, and 12%. The injections were effected through a hole made into the hollow of the fruit with a cork-drill. Alcohol was introduced in the amount of 1 cm^3 of solution for every 100 grams of the fruit. 11 green full-grown fruits were taken for each concentration of alcohol. The controls consisted likewise of 11 fruits.

The injected fruits ripened 5 to 8 days after the injection of the alcohol. The ripened fruits were distinguished by an intensified agreable odour of ether.

When 24 to 48% solutions were introduced, scalds was observed where the alcohol came in direct contact with the tissues. No differences in rapidity of ripening under injections of different concentrations were noted.

The control melons ripened 25 days after the beginning of the experiment.

The second experiment explained the effect of a temporary subjection of melons to anaerobic conditions on the rate of the ripening.

Exsiccators were used as chambers in which to obtain anaerobic conditions.

After being held in the exsiccators, the melons were brought to ripeness in ordinary room conditions. Four exsiccators were used in all. Two melons of the Biriuche-Kutskaya variety were placed in each of them.

The 1st exsiccator was opened 12 hours after charging, the 2nd after 24 hours, the 3rd after 36 hours, and the 4th after 48 hours.

The melons of the 1 st exsiccator ripened 12 days after charging: those of the 2nd after 8 days; those of the 3rd after 13 days and those of the 4th after 22 days. The control melons ripened in 14 days.

At the same time, a calculation was conducted of the dunamics of C_2H_5OH and CH_3COH when the melons were kept in anaerobic conditions (CH_3COH and C_2H_5OH were determined immediately after the removal of the fruits from the exsiccator. In the contol fruits CH_3COH and C_2H_5OH were determined at the moment of setting up the experiment). The results are shown in Table 3.

Table 3

Length of stay in exsiccator in hours	CH_3COH in mg per 100 grams of fruit	C_2H_5OH in mg per 100 grams of fruit	
22	0.110	22.0	The two last experiments show that the products of anaerobic metabolism (no matter whether they were introduced from outside or whether their formation was induced by anaerobic conditions) lead to a hasten of ripening.
24	0.80	41	
36	1.10	75.6	
48	1.93	80.0	
Control	0.088	18.0	

Gore ([22]) found that the fruits of the persimmon after being subjected to the action of chloroform and ether, lost their astringent taste (that is. in their qualities were similar to ripe fruits). Analogous results were obtained by Overholser ([23]), using alcohol vapour and an ethylene preparation. Soldatenkov ([24, 25, 26]) obtained a hastening of the ripening of tomato and persimmon fruits by alcohol vapours and a mixture of ethylene with alcohol. In 1935, working together with Bulatov and Stoljarov ([27]) (Gorki Agricultural Institute) we obtained a hastening of the ripening of tomato fruits by the action of ethylene, its preparations, temporary anaerobiosis, by injections of $CH_3COCOOH$, C_2H_5OH, CH_3COH and the action of warm baths.

As special investigations have shown (their results will be published in the near future), in all the cases enumerated there occurs a deflection of metabolism towards C_2H_5OH formation.

A consideration of the literary and experimental data lead to the conclusion that temporary anaerobiosis, as likewise all factors promoting the deflection of metabolism towards alcohol formation, hasten the ripening of fruits.

Consequently, the essence of the hastening of ripening may be regarded as a deflection of metabolism toward intensified anaerobic processes, or as a more rapid onset of the state whose rise in ordinary ripening is slow.

It is now perfectly clear that, being themselves the results of the natural development of fruits, the products of the anaerobic processes, including CH_3COH and C_2H_5OH, are factors hastening ripening, important links in the complicated chain of processes that we call r i p e n i n g. It is likewise clear that the sooner this link is formed, the sooner the ripening of fruits will take place.

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BOTANIK

BEOBACHTUNGEN ÜBER DIE VARIABILITÄT VON *SCENEDESMUS* MEYEN IN REINKULTUREN

Von S. N. JAGUZINSKIJ

(Vorgelegt von A. A. Richter, Mitglied der Akademie, d. 19. XI. 1935)

Im Jahre 1933 veröffentlichte O. V. Troickaja eine Arbeit ([4]), in welcher der Autor von neuen Gesichtspunkten an die Systematik der *Protococcales* herantritt und unter Heranziehung der experimentell-genetischen Methode zur Erforschung der systematischen Stellung jeder Form die Ansicht äussert, dass die gesamte Systematik der *Protococcales* umgearbeitet werden muss, da die bisher bei ihnen angewandte descriptiv-morphologische Methode infolge der starken Modifikations-Variabilität hier unbrauchbar ist. Unter 6 untersuchten Gattungen befindet sich auch die Gattung *Scenedesmus*, von welcher der Autor zwei Arten untersucht hat: *S. obliquus* und *S. quadricauda*. Obwohl die Untersuchungen von Troickaja in Bezug auf *S. quadricauda* nur orientierender Art waren, wurden doch interessante Resultate erzielt; insbesondere konstatierte der Autor in Bezug auf die für diese Art charakteristischen Auswüchse grosse Mannigfaltigkeit, und schreibt, dass dieses Merkmal bei diesem genotypischen Material sehr stark variiere, und folglich nicht die Bedeutung eines systematischen Merkmals haben könne.

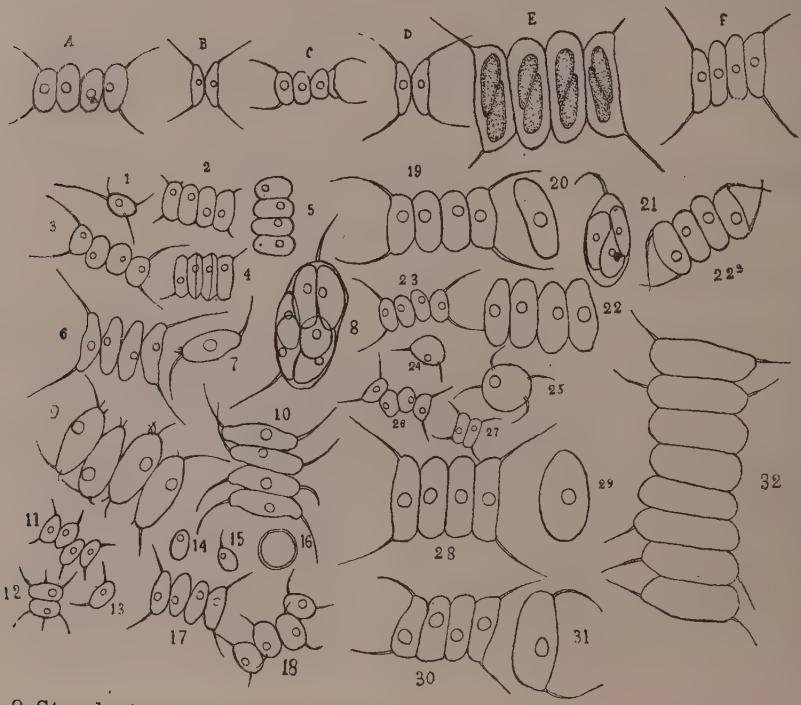
Die Untersuchung der Variabilität dieses, sowie nebenbei auch anderer systematischer Merkmale bildet den Gegenstand der vorliegenden Arbeit.

Material und Methodik

Als Material für die Arbeit dienten vierhörnige *Scenedesmus* aus dem Fluss Kljasma bei der Bolshevo- Biologischen Station. Die Abteilung der Klonen wurde auf dem Wege der Aussaat einzelner mittels der vereinfachten Micropipette Hollerbach's ([3]) isolierter Individuen in einen Mineralnährboden bewerkstelligt. Dieses Isolierungsverfahren ist zweifellos bedeutend leichter und exakter als die von Troickaja angewandte Methode der Isolierung; ausserdem ermöglicht sie die Erzielung einer von anderen Algen freien Kultur sogar in dem Falle, wenn die zu untersuchende Art in vereinzelten Exemplaren vorliegt.

Nachdem die Entwickelung einer genügenden Zahl von Cönobien stattgefunden hatte, wurden diese in Nährböden von Uspenskij, Benecke, Bei-

jerink, № 1*, in Agar nach Beijerink ausgesät. In manchen Fällen wurden auch andere Nährböden benutzt: nach Knopp (1/4), Sachs (1/3), Agar auf Nährboden № 1, Nährboden nach Beijerink+0.5% Glycose. Alle Nährböden wurden durch strömenden Dampf sterilisiert, ebenso wie die zur Entnahme des Materials dienenden Pipetten. Während der Winterzeit wurde für ergänzende Beleuchtung mit einer elektrischen Lampe von 75 W (täg-



lich 8 Stunden) gesorgt; im Sommer verläuft das Wachstum der Kulturen auch bei Sonnenlicht gut.

Alle Zeichnungen und Messungen wurden mit Reichert's Mikroskop, Apochromat (2 mm), Kompensationsokular 6 und Zeichenapparat nach Abbé ausgeführt. Fast alle Zeichnungen sind von 20—40 Tage alten Kulturen gemacht.

Untersucht wurden bisher** nur 6 Klonen. Die Resultate sind im Nachfolgenden dargelegt.

Beschreibung der Resultate

Ich gebe hier nur eine kurze Beschreibung jeder Klone, indem ich ausführlichere Mitteilungen für eine weitere Arbeit vorbehalte, die auf Grund umfassenderer Untersuchungen veröffentlicht werden wird.

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* Nährboden № 1 hat folgende Zusammensetzung: H_2O—1000; KNO_3—0.5; $MgSO_4$—0.1; Na_4HPO_4—0.2; FeCl—Spuren.

** Juli 1934—März 1935.

Klone 1. Erhalten durch Aussaat der auf Abb. *A* dargestellten, nach Brunnthaler ([1,2]) als *S. quadricauda* zu bestimmenden Kolonie. Die Länge der Randzellen dieser Kolonie beträgt 11 μ. In der Kultur schwankt die Zellenlänge zwischen 4.5 und 14.6 μ. Mittlere Länge in Uspenskij's Nährboden 8.7 ± 0.0989 μ. Zellenform schmal elliptisch mit fast zugespitzten Enden bis gerundet elliptisch. Randzellen oft vor dem Ende mit einer Falte. Cönobien meist 4-zellig (70—90% der Individuen), es kommen jedoch 8- und 2-zellige und vereinzelte Zellen vor. Die Zahl der Stacheln schwankt zwischen 4 und 0. An den mittleren Zellen entwickeln sich keine Stacheln. Einzelzellen gewöhnlich mit 3 und 4 Stacheln. In Benecke's Nährboden zeigen sich nicht selten Individuen mit längsgerippten Mittelzellen (vgl. Abb. 1—5). Die Cönobien entwickeln sich mit den Stacheln miteinander verkettet und bilden ein flockenartiges Sediment.

Klone 2. Erhalten durch Aussaat der auf Abb. *B* dargestellten Kolonie mit einer Zellenlänge von 9 μ. Nach der Zellenform nähert sie sich mehr *S. opoliensis* P. Richter, aber ihre Dimensionen entsprechen nicht der Beschreibung. Zellenlänge in der Kultur 5.5—14.6 μ. Mittlere Länge in Nährboden Uspenskij's 7.97 ± 0.0207 μ. Zellen mit abgerundeten Enden, von schmal ovaler bis kugeliger Form. Keine Falten vor den Enden der Randzellen, Cönobien meist zweizellig (bis 60%). Viele einzellige Individuen (bis 100% in Agar). Zwei- und vierzellige Individuen mit nicht weniger als 4 Stacheln, deren Anordnung aus Abb. 11—18 ersichtlich ist. Einzellige Individuen mit 1 oder 2 Stacheln oder ohne solche; offenbar handelt es sich um zerfallene Cönobien. Die Cönobien entwickeln sich einzeln und bilden ein Sediment am Boden des Gefässes.

Klone 3. Erhalten durch Aussaat der auf Abb. *C* dargestellten vierzelligen Kolonie mit einer umgekommenen Zelle. Zellenlänge in dieser Kolonie 7 μ. *S. quadricauda* (Turp.) Bréb. näher stehend. Zellenlänge in der Kultur 5.5 bis 12.7 μ. Mittlere Länge im Nährboden Uspenskij's 7.08 ± 0.0233 μ. Zellenform schmal oval bis kugelig; Randzellen zuweilen mit einer Falte vor der Spitze. Die Enden der Zellen abgerundet. Cönobien grösstenteils zweizellig (70—90%). Nur die Randzellen mit Stacheln versehen, je zwei an einer Zelle. Einzelne Zellen ohne oder mit 2—4 Stacheln (besonders in Beijerink's Nährboden). Cönobien schwach mit ihren Stacheln verkettet, ein Sediment am Boden des Gefässes bildend (Abb. 23—27).

Klone 4. Erhalten durch Aussaat einer zweizelligen Kolonie mit einer Zellenlänge von 13.5 μ. Nach der Zellenform der Beschreibung von *S. opoliensis* P. Richter entsprechend, aber etwas kleiner (vgl. Abb. *D*). Zellenlänge in der Kultur 13.5—28.5 μ. Mittlere Länge im Nährboden Uspenskij's 15.34 ± 0.1073 μ. Zellen spindelförmig mit etwas aufgetriebener Mitte, die Enden abgerundet; vor den Enden der Randzellen, bei schmalen Zellen, deutliche Falten. Cönobien aus 2—8 Zellen, meist zweizellig (80—90% der Individuen); auch einzelne Zellen kommen vor. Randzellen der Cönobien stets mit Stacheln; in Beijerink's und in № 1 Nährböden auch die mittleren Zellen mit solchen. Ausserdem entwickeln sich in Beijerink's Nährboden an den Zellenenden dünne Borsten (nicht bei allen Zellen).

Einzelzellen mit 1—2 Stacheln oder ohne. Die Cönobien entwickeln sich einzeln, und bilden ein Sediment (Abb. 6—10).

Klone 5. Erhalten durch Aussaat einer grossen in Vermehrung begriffenen Kolonie mit einer Zellenlänge von 22 μ. Nach dem Bau der Zellen und den Dimensionen entspricht die Kolonie der Beschreibung von *S. quadricauda* (Turp.) Bréb., mit dem Unterschied jedoch, dass die Stacheln an den Enden jeder Randzelle von ungleicher Länge sind (vgl. Abb. *E*).

Zellenlänge in der Kultur 11.6—23.3 μ. Mittlere Länge im Nährboden Uspenskij's 12.99±0.186 μ. Zellen lang elliptisch, mit abgeflachten Berührungsflächen und breit abgerundeten Enden. Die Enden der Randzellen nicht selten zur Basis des Stachels etwas herangezogen; dann bildet sich vor den Enden eine Falte. Cönobien aus 8, 2 oder 4 Zellen bestehend, meist aus 4 (50—70% der Individuen). Stacheln nur an den Randzellen oder nicht vorhanden (in Beijerink's Nährboden). Einzelzellen ohne oder mit 1—2 Stacheln. Die Cönobien entwickeln sich einzeln und bilden ein Sediment (Abb. 19—22).

Klone 6. Erhalten durch Aussaat einer vierzelligen Kolonie mit einer Zellenlänge von 15.5 μ, deren Bau der Beschreibung von *S. quadricauda* (Turp.) Bréb. entsprechend (vgl. Abb. *F*). Zellenlänge in der Kultur 12—22.5 μ. Mittlere Länge im Nährboden Uspenskij's 15.36±0.157 μ. Zellenform ebenso wie in der Klone 5, nur sind die Randzellen öfters birnförmig. Cönobien aus 8.2 oder meist 4 Zellen bestehend (80—90% der Individuen). Endzellen stets mit je 2 Stacheln. Im Nährboden № 1 mit Glycose entwickeln sich Cönobien, die auch an den mittleren Zellen Stacheln aufweisen. Einzelzellen ohne oder mit 2 Stacheln. Die Cönobien entwickeln sich einzeln und bleiben in jungen Kulturen (bis 2 Monate) sämtlich suspendiert (Abb. 28—32).

Fassen wir diese Beobachtungen zusammen, so können wir Folgendes feststellen.

1) Die Länge der Zellen erweist sich als eine für die Klone recht charakteristische Eigenschaft und kann als systematisches Merkmal dienen.

2) Die Umrisse der Zellen sind so variabel, dass sie als systematisches Merkmal kaum Verwendung finden können.

3) Die Zahl und Lage der Stacheln ist in manchen Klonen recht variabel, aber die Richtung dieser Variabilität ist in jeder Klone eine bestimmte, das heisst entweder im Sinne einer Abnahme oder einer Zunahme der Stachelzahl.

4) Die Zellenzahl in den Cönobien verändert sich stark, aber wiederum in verschiedenen Klonen verschieden.

5) Die Bildung einzelner Zellen in verschiedenen Klonen geht verschieden vor sich. In einigen Fällen geschieht das auf dem Wege des Zerfalls der Cönobien, in anderen durch Formierung einzelner Zellen innerhalb der Mutterzelle (dafür spricht das Vorhandensein einzelner Zellen mit 4 Stacheln in den Klonen 1 und 3).

Wenn man die von uns erhaltenen Resultate mit den Verallgemeinerungen vergleicht, die Troickaja (4) macht, so kommt man zum Schluss, dass nicht alle unsere Resultate in den Rahmen dieser Verallgemeinerungen passen, und dass sie in manchen Fällen den letzteren sogar widersprechen.

Ich will hier auf die wesentlichen Widersprüche hinweisen.

1. Man kann nicht den Annahmen von Troickaja zustimmen, dass man «nur in Bezug auf die soeben aus der Mutterhülle herausgetretenen jungen Zellen die Gewissheit haben kann, dass dieselben die dieser Art eigenen Umrisse besitzen.» Zweifellos verändert sich die Zellenform mit dem Alter, besonders infolge des Fehlens einer Vermehrung oder eines Wachstums der jungen Zellen im Inneren der Mutterzelle (Abb. 8), allein ein solches Merkmal wie die Zahl und Lage der Stachel erweist sich als vom Alter unabhängig, während Milieu und Erbfaktoren auf dasselbe einen sehr starken Einfluss ausüben. Infolgedessen finde ich, dass im gegebenen Fall von einer der betreffenden Art eigentümlichen Form nicht die Rede sein kann, sondern dass man zur Charakteristik irgendeiner systematischen Einheit die Gesamt-

heit der Veränderungen heranziehen muss. Das muss umso eher betont werden, als auch Troickaja selbst, drei Zeilen weiterhin, schreibt: «die Algenzellen, die aus den veränderten Zellen heraustreten, können sich unter analogen äusseren Verhältnissen ebenso als verändert erweisen.»

2) Die von Troickaja angezeigte allgemeine Richtung, in welcher die Veränderungen der Zellenform «wahrscheinlich bei allen Gattungen der *Protococcales*» vor sich gehen, ist wenig auf die von mir erforschten *Scenedesmus* anwendbar.

a) Die Länge der «Riesenzellen» übertrifft die mittlere Länge nicht mehr als um das doppelte.

b) Die Zahl und die Länge der Stachel hängt nicht von den Dimensionen der Zellen ab.

c) Die Verbindung zwischen den Zellen der Cönobien wird nicht nur infolge der Abrundung der Zellen, sondern auch unabhängig von ihr unterbrochen.

3. Die von Troickaja erzielten Resultate führten sie zur Folgerung, dass in Bezug auf die von ihr untersuchten Merkmale (Form und Grösse der Zellen, Länge und Zahl der Auswüchse, Zellenzahl in den Cönobien) die *Protococcales* keinen systematischen Polymorphismus besitzen (im Sinne der Zusammensetzung jedes Linneons aus einer grossen Zahl sich durch geringe Merkmale unterscheidender Arten).

Die Beobachtungen an *Scenedesmus* haben das Umgekehrte gezeigt. Jede von den sechs untersuchten Klonen besitzt eine charakteristische Summe von Merkmalen, jede reagiert auf ihre eigene Art auf Veränderungen des Mediums. Wenn es einstweilen noch verfrüht ist, über dieselben als über selbstständige Arten zu sprechen, so scheint mir doch, dass man sie als bestimmte systematische Einheiten niederer Kategorie als Arten ansehen kann.

Die weitere Erforschung neuer Klonen wird uns die Möglichkeit geben, darüber zu urteilen, und die Verbreitung des ohne Zweifel bestehenden Polymorphismus der Arten in dieser *Scenedesmus*-Gruppe zeigen.

Biologische Station
Bolschewo.

Eingegangen
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ZITIERTE LITERATUR

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ZOOLOGY

THE FAUNA OF AVACHA BAY AND ITS DISTRIBUTION INTO COMMUNITIES

By A. M. POPOV

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Avacha Bay is one of the most interesting localities along the eastern coast of Kamchatka. It penetrates deeply into the peninsula, having a length of 15 miles and a width of 12 miles. It is connected with the sea by a narrow mouth about 1 mile wide. Since there are hardly any data on the hydrobiology of the eastern coast of Kamchatka, and since, as regards Avacha Bay, only the scanty (already antiquated) data of Grebnitsky are available, the hydrobiological material collected by me in 1930 is of considerable both theoretical and practical interest. During the summer of that year I succeeded in collecting a great deal of material on the hydrobiology and ichthyology of this bay, based on a large number of stations (about 100). In this short paper I communicate in brief the preliminary data only.

Avacha Bay, because its isolation and a number of rivers emptying into it, as the Paratunka and Avacha, plays an important role in the life of many fishing and non-fishing districts of Kamchatka. The fauna of the bay is of a subarctic character with a small admixture of boreal forms, on the one hand (mainly near the mouth of the bay), and of arctic ones, on the other. Most of the forms are common both for the Okhotsk and Bering seas. The population is very scarce in comparison with that of the open sea. The following flora is characteristic of Avacha Bay: fucoid seaweeds, *Entoromorpha*, sea-grass *(Zostera pacifica)*, *Laminaria* and a number of others, as *Litothamnion*, etc. Representatives of almost all the principal groups of fauna are to be found here, but the principal marine elements are noticeably less abundant. So, for instance, the *Porifera*, *Actinia*, *Echinodermata* and higher crustaceans *(Decapoda)* are scarce as regards their qualitative composition. The prevailing groups are *Polychaeta*, *Mollusca*, *Isopoda* (several forms), *Amphipoda* (several forms), and *Pisces* (in very great number).

The horizontal und vertical distribution of the fauna in Avacha Bay depends both on the character of the region and on the character of the bottom. As for the former, the bay may be divided into the following areas markedly differing in the composition of their fauna and in their general physico-geographical conditions: 1) the mouth of Avacha Bay and the adjacent region, 2) the central part of the bay, 3) the head of the bay, and 4) Rakovyi and Taryinski inlets.

1. In the first area we find a luxuriant growth of *Laminaria* and *Fucus*. The bottom is for the most part rocky. The fauna is rich in forms common in the open sea, but absent from the bay. The following are to be found here: *Alcyonaria*, flat sea urchins, great quantities of shrimps and large *Decapoda*, *Paralithodes brevipes* and, lastly, *Cephalopoda*.

2. The central area is characterized by the abundance of mud, growths of fucus near the shores, rare growths of laminaria. An abundance of *Polychaeta* of the family *Maldanidae* on mud and of those with calcareous tubes, of the family *Serpulidae*, near rocky shores is to be noted. There is a number of forms of different marine groups of animals which occur here in great quantities (sponges, starfishes, sea-anemones, sea urchins, isopods, some forms of gastropods, ascidians and others).

3. The third area is characterized by a sandy bottom, considerable shallowness and the development of *Entoromorpha* growths (in the north-western corner) and of fresh-water plants (in the north-eastern corner). The fauna consists of but single forms, occurring here in enormous quantities (*Tellina baltica*, *Platichthys stellatus*, and others).

4. Rakovyi and Taryinski inlets. Oozy mud. Growths of sea-grass (*Zostera*) both in the head of the inlets and in their lagoons. Great development of individual representatives of separate groups (*Idathea tricuspidata*, *Littorina* sp.). In the very ends of the inlets (Rakovyi), arctic relict forms are to be found, e. g. *Chiridothea entomon* and others.

The vertical distribution of the fauna closely depends upon the depths and the bottom. The following bottoms exist in the bay: rock, pebble, sand and mud.

The Littoral Zone is variously developed in the various parts of the coast. The littoral may be rocky, sandy, or muddy. Its width depends on the character of the shore, being less in steep and rocky shores and greater in gently sloping shores. Littorals with rocks and cliffs (e. g., the shore of the mouth of Avacha Bay, many localities of the Taryinski peninsula, Khlebalkan island, and others) have a strongly developed fucus zone. The population of these fucus growths consists of molluscs: *Littorina* sp., *Ballanus porcatus*, *Idathea tricuspidata*, *Opisthocentrus ocellatus*, *Pholis dolichogaster*, *Alectrials alectrolophus*, and others. The fucus and its population descend much below the level of the low water mark and part of them are also to be found in the zone of high water mark.

On the sandy parts of the littoral (as, for instance, in some localities of the Taryinski peninsula, and others) there exist mounds of laminaria cast up by the sea and containing large quantities of crustaceans.

Outside of the mounds, these crustaceans bore small holes in which they find shelter till the next high tide.

The Sublittoral. Below the tidal zone the development of the fauna depends on the character of the adjoining shore, the depth and the bottom. On a pebbly bottom (as in some localities of the eastern shore of Rakovyi inlet) the population consists of large molluscs *Littorina*, hermit crab, *Pagurus pubescens*, and others. In some localities of this zone there occur accumulations of *Mytilus*. It should be noted that some rocks covered with *Littorina* emerge during low tide.

At the head of the Bay, on the sandy foreshore, there are scattered small patches of *Entoromorpha*. The fauna found here is not always associated with this seaweed. Thus, *Tellina baltica*, *Buccinum* sp., *Scolaphas armiger*, *Gammarus locusta* and *Platichthys stellatus* are not associated with it. There occur in some localities the fish *Hemitripterus americanus* and the

crab *Paralithodes latypus*. On rocky bottom, near steep shores (as in the mouth of the Rakovyi inlet and near the Gremuchi stream), there occur at a depth of several metres enormous quantities of the calcareous tubes of *Serpulidae*. Here we have an exceptionally rich fauna consisting of sponges, *Bryozoa*, sessile ascidians, hydroids, one ophiuran, crustaceans, as *Caprellidae* in the sponges, molluscs and fishes. Among the latter we have: *Ceratocottus diceraus*, *Hypsagonus quadricornis*, *Artediellus pacificus ochotensis*, and others. At the head of Rakovyi and Taryinski inlets we find *Zostera* distributed from the topmost horizons of the sublittoral down to a depth of several metres. The population is very abundant and consists of sedentary forms of molluscs, crustaceans *Idothea tricuspidata*. The following fishes are characteristic of this horizon: *Pallasina barbata*, *Pholis pictus*, *Gymnocanthus pistilliger*, *Myoxocephalos stelleri*, *Liopsetta pinifasciata*, etc.

There occur extensive growths of laminaria in the central area of the bay and its mouth at a depth of several metres. This seaweed is especially abundant in the district of the Toporkov Island, and generally along the coast to the south of Avacha Bay, while it is most probably absent to the north of the mouth of the Avacha. The fauna is very varied and abundant. Here we find the tubes of *Spirorbis* on thalli, hydroids, sponges covering rhizoids, ascidians. On thalli we find very large quantities of molluscs, crustaceans, as *Gammarus locusta*, shrimps, crabs, as *Paralithodes brevipes*, flat sea urchins, cephalopods, etc. The following fishes inhabit this area: *Histiocottus bilobus*, *Blepsias cirrhosus*, *Nautichthys pribilovius*, *Gymnocanthus pistilliger*, *Podothecus gilberti*, *Occadodecaedron*, *Cyclopterichthys ventricosus*, *Eumicrotremus orbis*, *Hexagrammas stelleri*, *Triglops pingeli*, *Limanda aspera*, etc. The oosy mud of the central area of the bay is inhabited by great quantities of *Polychaeta* of the family *Maldanidae*, as mud tubes of various sizes. Other repsesentatives of the fauna found here are common along this coast.

Of fishes, the flat-fish and, it seems, the *Gadidae* inhabit this locality.

Thus, we see that the fauna of Avacha Bay has both a vertical distribution and a distribution into communities. Of these latter the following may be mentioned: 1) the fucus community, 2) the community of *Laminaria* mounds, 3) the pebble community with *Pagurus pubescens* (e. g. the community of rocks with *Littorina*, the community of *Mytilus* sp.), 4) the *Entoromorpha* community, 5) the community of rocks with calcareous tubes of *Serpulidae*, 6) the *Laminaria* community, 7) the *Zostera* community, 8) the community of *Maldanidae* mud.

As had been mentioned above, the distribution of these communities in the bay depends on the characteristics of the district. Thus, on the littoral of the mouth the fucus community and that of *Laminaria* mounds are prevalent, while in the sublittoral the *Laminaria* community predominates. The same communities are observed to occur on the littoral, but on the sublittoral two new communities occur, viz., rocks with tubes of *Serpulidae* and *Maldanidae* mud. Both the fucus community and that of the *Laminaria* mounds are almost absent from the north part of Avacha Bay. In the sublittoral, the above-mentioned communities disappear and are replaced by the *Entoromorpha* community. Lastly, there are the fucus community on the littoral of the inlets and the *Zostera* community on the sublittoral.

It is interesting to note that besides the above elements, closely associated with the different communities, there also exists in the bay a number

of forms having a very widespread, almost universal, distribution. Such are the *Actinia* sp. of very large dimensions, large red starfishes *Asteriae* sp., the sea crab *Heirogonus heirogonus*, and the sea urchin *Strongylocentratus drobachiensis*.

The water of the Petropavlovsk Harbour is somewhat polluted, so the fauna there is much scantier, with the exception of fishes and the above-mentioned widely distributed forms.

In conclusion it is necessary to emphasize the importance of further, more detailed biological investigations of the fauna of the bay and of hydrobiological investigations along the eastern coast of Kamchatka.

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ZOOGEOGRAPHICAL CONCLUSIONS FROM AN EXPEDITION TO THE MIDDLE SIKHOTE-ALIN RANGE

By A. I. KURENCOV

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Up to the present time the fauna of the Sikhote-Alin range remains insufficiently studied. My own investigations of it have been carried on throughout the last few years.

In the present communication the author gives his principal conclusions with respect to the zoogeography of the middle districts of this mountainous region as concerns only the *Lepidoptera*, which were the object of his special investigations during the zoological expedition dispatched by the Far Eastern Branch of the USSR Academy of Sciences in 1934.

The region extending from town Iman to the principal chain of the Sikhote-Alin in the East (the route of the expedition, approximately 400 km), presents a characteristic latitudinal profile of the Ussurian fauna. In this stretch we observe not only ecological differences of fauna, manifesting themselves in a variety of conditions of animal habitants, we also note, nearer to the principal range of the Sikhote-Alin (in the basin of the Kolumbe River), well expressed zoogeographical boundaries of the separate elements of the fauna.

Proceeding from West to East along the valley of the Iman, along with the growing character of the mountainous relief and the impoverishment of flora of the mixed taiga of Manchurian type, there is a gradual fall in the wealth of the fauna of these forests. Together with this, there is more and more manifested a biocenetic reorganization of the latter with species of coniferous forests.

In the Iman district there can be distinguished in the fauna of the mixed Manchurian woods: 1) species covering the whole district and reaching across the Sikhote-Alin to the coast (*Limenitis populi ussuriensis* Stg., *Angerona prunaria* L., *Abraxas grossulariata* L., *Acidalia nemoraria* Hb. *Cepphis advenaria* Hbp., and others); 2) species found on the East to the lower reaches of the Kolumbe River (*Neptis speyeri* Stg., *N. raddei* Brem., *Limenitis camilla japonica* Mén., *Actias selene mandschurica* Stg., *Boarmia stipitaria* Ob., and others); 3) species penetrating still farther to the East, to the middle course of the Kolumbe River (*Papilio maackii* Mén., *Apatura schrenckii* Mén., *Lethe epimenides* Mén., *Neptis philyra* Mén., *Limenitis amphissa* Mén., *Problepsis phoebearia* Ersch., and many others); 4) species reaching westward to the uppermost courses of the Kolumbe River but not acrossing the range to the coast (*Aporia hippia* Brem., *Ar-*